COMPTES RENDUS

(DOKLADY)

DE L'ACADÉMIE DES SCIENCES DE L'URSS

NOUVELLE SÉRIE

1945

VOLUME XLVII

№ 8

ÉDITION DE L'ACADÉMIE DES SCIENCES DE L'URSS
MOSCOU

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NOUVELLE SÉRIE

13-me année

Paraissant tous les dix jours

1945

VOLUME XLVII, № 8

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Édition de l'Académie des Sciences de l'URSS
Moscou 1945

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

MATHÉMATIQUES

SUR L'IMMERSION D'UN ESPACE A CONNEXION AFFINE DANS UN ESPACE AFFINE

Par G. LAPTEV

(Présenté par N. N. Lusin, de l'Académie, le 8. VIII. 1944)

Considérons dans un espace affine à N dimensions une surface S_n à n dimensions ($n < N$). Supposons que les coordonnées d'un point M qui parcourt cette surface sont des fonctions analytiques de n paramètres $u^1, \ldots, u^n$. Désignons respectivement par $\mathfrak{S}_n(u)$ et $\mathfrak{N}_{N-n}(u)$ le plan tangent et la normale à la surface S_n au point $M(u)$. D'ailleurs, on nommera normale $\mathfrak{N}_{N-n}(u)$ un plan arbitraire à $N-n$ dimensions qui possède un seul point commun $M(u)$ avec le plan tangent $\mathfrak{S}_n(u)$, et qui est défini au moyen de paramètres qui sont des fonctions analytiques de $u^1, \ldots, u^n$.

La partie principale de la représentation de $\mathfrak{S}_n(u+du)$ sur $\mathfrak{S}_n(u)$ obtenue en réalisant une projection parallèlement à $\mathfrak{N}_{N-n}(u)$ établit sur S_n une connexion affine sans torsion. En variant l'ensemble des normales $\mathfrak{N}_{N-n}(u)$, on peut obtenir sur S_n une infinité de connexions affines, établies au moyen de projections (1).

M. Cartan (2) a posé le problème d'immersion d'un espace à connexion projective dans un espace projectif; d'une manière analogue, on peut poser le problème suivant sur l'immersion d'un espace à connexion affine dans un espace affine.

Soit A_n un espace affine à n dimensions. Il s'agit de trouver dans un espace affine $\mathfrak{A}_N$ à N dimensions, une surface S_n à n dimensions, et de faire correspondre à chaque point de sette surface une normale $\mathfrak{N}_{N-n}$ à $N-n$ dimensions de telle manière que la connexion affine établie sur S_n, en réalisant la projection au moyen de cette famille de normales $\mathfrak{N}_{N-n}$, coïncide avec la connexion de l'espace A_n.

Pour que la solution de ce problème soit possible, il est nécessaire que l'espace A_n que l'on plonge ne possède pas de torsion. Si cette condition est vérifiée, les formes de Pfaff ω^i, ω^i_k, qui déterminent la connexion dans l'espace A_n, seront liées par les équations de structure suivantes

$$(\omega^i)' = [\omega^k \omega^i_k], \quad (\omega^i_k)' = [\omega^l_k \omega^i_l] + R^i_{kpq}[\omega^p \omega^q], \quad i, k, l = 1, \ldots, n \qquad (1)$$

où R^i_{kpq} sont les composantes du tenseur de torsion, liées par les identités de Bianchi:

$$R^i_{k(pq)} = 0, \quad R^i_{kpq} = 0$$

$$[dR^i_{kpq} - R^i_{lpq}\omega^l_k - R^i_{klq}\omega^l_p - R_{kpl}\omega^l_q + R^l_{kpq}\omega^i_l;\ \omega^p;\ \omega^q] = 0 \qquad (2)$$

D'ailleurs, on désigne par $'$ les dérivées extérieures, et par [ ] les produits extérieurs.

En désignant par ϑ^k, ϑ^J_K les formes de Pfaff qui déterminent le déplacement infinitésimal du repère mobile dans l'espace affine $\mathfrak{A}_N$ à N dimensions, nous aurons pour ces formes les équations de structure suivantes:

$$(\vartheta^K)' = [\vartheta^L \vartheta^K_L], \quad (\vartheta^J_K)' = [\vartheta^L_K \vartheta^J_L], \quad J, K, L = 1, \ldots, N \tag{3}$$

Nous supposerons que le repère mobile dans $\mathfrak{A}_N$ est choisi de telle manière que pour un point M, situé sur S_n, les vecteurs $\vec{e}_1, \ldots, \vec{e}_n$ sont situés dans le plan $\mathfrak{S}_n$ tangent à S_n au point M. Alors, on peut donner la forme suivante au système d'équations différentielles de l'immersion et aux conditions de leur intégrabilité:

$$\vartheta^k - \omega^k; \quad \vartheta^\alpha = 0; \quad \vartheta^i_k = \omega^i_k; \quad [\omega^k \vartheta^\alpha_k] = 0;$$
$$[\vartheta^\alpha_k \vartheta^i] = R^i_{kpq} [\omega^p \omega^q]; \quad i, k, p, q = 1, \ldots, n; \quad \alpha, \beta = n+1, \ldots, N \tag{4}$$

Le problème d'immersion se réduit à l'intégration de ce système où l'on considère comme variables indépendantes les n coordonnées $u^1, \ldots, u^n$ du point mobile de l'espace A_n que l'on immerge, et comme fonctions inconnues les N^2+N paramètres qui déterminent la position du repère mobile $M, \vec{e}_1, \ldots, \vec{e}_N$ dans l'espace affine $\mathfrak{A}_N$.

Comme les formes ϑ^β_α ne figurent pas dans le système (4), on peut démontrer que le nombre des fonctions inconnues peut être abaissé jusqu'à $N+n-(N-n)^2$ par un changement convenable des variables. D'ailleurs, les formes ϑ^k, ϑ^i_k, ϑ^α_k, ϑ^α qui figurent dans le système (4) seront linéairement indépendantes entre elles par rapport aux différentielles des fonctions inconnues nouvelles.

Pour établir la possibilité de l'immersion j'étudie dans la présente Note la compatibilité du système (4) par la méthode de Cartan [3] appliquée pour la solution du problème sur l'immersion de l'espace riemannien [4] et de l'espace à connexion projective [5]. Cette étude se réduit à la construction d'une suite régulière d'éléments intégraux $E_1 < E_2 < \ldots < E_n$ dont le dernier élément E_n est un élément à n dimensions, sur lequel toutes les formes $\omega^1, \ldots, \omega^n$ sont linéairement indépendantes.

Comme les formes $\omega^1, \ldots, \omega^n$ peuvent subir une transformation linéaire quelconque, on peut, sans restreindre la généralité, déterminer chaque élément E_r à r dimensions de cette suite par le système suivant:

$$\begin{aligned} &\omega^{r+1} = \ldots = \omega^n = 0 \\ &\vartheta^\alpha_k = x^\alpha_{k1}\omega^1 + \ldots + x^\alpha_{kr}\omega^r \\ &\vartheta^i_\alpha = x^i_{\alpha 1}\omega^1 + \ldots + x^i_{\alpha r}\omega^r \\ &\vartheta^\alpha = 0; \quad \vartheta^i_k = \omega^i_k \end{aligned} \tag{5}$$

où les coefficients x^α_{kp}, $x^i_{\alpha p}$ doivent vérifier les relations suivantes qui résultent des équations (4) en tenant compte des relations (2):

$$x^\alpha_{qp} x^i_{\alpha r} - x^\alpha_{qr} x^i_{\alpha p} = 2R^i_{qpr}$$
$$x^\alpha_{\lambda p} x^i_{\alpha r} - x^\alpha_{\lambda r} x^i_{\alpha p} = 2R^i_{\lambda pr}$$
$$i \leqslant p < q < r; \quad r \leqslant \lambda \leqslant n; \quad 1 \leqslant i \leqslant n$$

Suivant Cartan [4] considérons x^α_{ki} et $x^i_{\alpha k}$ comme projections de certains vecteurs $\vec{x}_{ki}$ et $\vec{x}^i_k$ sur les axes d'un système rectangulaire des coordonnées

dans l'espace euclidien à $\nu = N - n$ dimensions. Le dernier système peut alors être présenté dans la forme suivante:

$$\left.\begin{aligned} \vec{x}_{qp}\vec{x}^i_r - \vec{x}_{qr}\vec{x}^i_p &= 2\vec{R}^i_{qpr} \\ \vec{x}_{\lambda p}\vec{x}^i_r - \vec{x}_{\lambda r}\vec{x}^i_p &= 2R^i_{\lambda pr} \end{aligned}\right\} E_r \qquad (6)$$

$$1 \leqslant p \leqslant q < r; \quad r \leqslant \lambda \leqslant n; \quad 1 \leqslant i \leqslant n$$

où $\vec{x}_{qp}\vec{x}^i_r$, $\vec{x}_{qr}\vec{x}^i_p$, *etc.* désignent les produits scalaires de vecteurs correspondants.

La démonstration de la possibilité de construire une suite régulière d'éléments intégraux se réduit à la démonstration de la résolubilité du système (6) par rapport aux vecteurs inconnus $\vec{x}^i_r$, $\vec{x}_{\lambda r}$ pour chaque valeur de r, et pour des vecteurs arbitraires $\vec{x}_{qp}$, $\vec{x}^i_{qr}$, $\vec{x}^i_p$, $\vec{x}_{\lambda p}$ successivement définis par le système (6) pour les valeurs précédentes de l'indice r.

La condition nécessaire et suffisante pour que le nombre des équations du système (6) soit pour chaque valeur $r = 2, \ldots, n$ au moins égal au nombre des inconnues, est donnée par l'inégalité

$$N \geqslant {}^1/_2\,(n^2 + 2n - 1)$$

Si cette condition est vérifiée, on peut démontrer la résolubilité de la suite des systèmes (6) dans le sens indiqué. A cet effet, on démontre d'abord le lemme suivant.

Lemme 1. Soit r_0 le nombre entier défini par les inégalités ${}^1/_2\,n - 1 < r_0 \leqslant {}^1/_2\,n$. Alors pour la suite des systèmes $E_2, E_3, \ldots, E_{r_0}$ que l'on obtient du système (6) pour $r = 2, 3, \ldots, r_0$, on peut prendre un système de vecteurs $\vec{x}^i_1$; $\vec{x}_{i1}$ $(i = 1, \ldots, n)$, et déterminer successivement les systèmes de vecteurs $\vec{x}^i_2$, $\vec{x}_{\lambda_2}$ $(\lambda_2 = 2, \ldots, n)$; $\ldots$; $\vec{x}^i_{r_0}$, $\vec{x}_{\lambda_{r_0} r_0}$ $(\lambda_{r_0} = r_0, \ldots, n)$ de telle manière que les conditions suivantes soient vérifiées: 1) tous les systèmes de la suite $E_2, E_3, \ldots, E_{r_0}$ sont vérifiés; 2) chaque système de la suite est formé d'équations linéairement indépendantes pourvu que tous les systèmes précédents soient déjà vérifiés; 3) le nombre des dimensions de l'espace vectoriel $[\vec{x}_{i1}; \vec{x}_{\lambda_2 2}; \ldots; \vec{x}_{\lambda_{r_0} r_0}; N(\vec{x}^i_1, \ldots, \vec{x}^i_{r_0})]$, construit sur les vecteurs $\vec{x}_{i1}, \vec{x}_{\lambda_2 2}; \ldots; \vec{x}_{\lambda_{r_0} r_0}$ et les vecteurs perpendiculaires aux vecteurs $\vec{x}^i_1, \ldots, \vec{x}^i_{r_0}$, est égal à $N - n - 1/2\, r_0\,(r_0 - 1)$.

Chaque suite de solutions assez proche à celle qui a été indiquée, jouit aussi de toutes les propriétés énoncées.

En tenant compte du lemme 1, on peut démontrer le lemme suivant.

Lemme 2. En prolongeant la suite des solutions obtenues dans le lemme 1, choisissons le vecteur $\vec{x}_{n,\, r_0+1}$ et $n - r_0 - 1$ vecteurs linéairement indépendants $\vec{x}_{r_0+1,\, r_0+1}; \ldots; \vec{x}_{n-1,\, r_0+1}$ de telle manière que les vecteurs $\vec{x}_{n1}; \ldots; \vec{x}_{n,\, r_0-1}$ précédemment définis soient linéairement dépendants des vecteurs $\vec{x}_{qp}$, $\vec{x}_{\lambda p}$, $\vec{x}_{\lambda r_0+1}$ $(1 \leqslant p \leqslant \leqslant q \leqslant r_0; \quad r_0 < \lambda < n)$, et que le nombre des dimensions de l'espace $[\vec{x}_{qp}, \vec{x}_{\lambda p}, \vec{x}_{ns}, N(\vec{x}^i_1, \ldots, \vec{x}^i_{r_0-1})]$ $(s = 1, \ldots, r_0 + 1; \; i = 1, \ldots, n)$ soit au moins égal au plus petit de deux nombres $N - n$ et $N - {}^1/_2\, r_0\,(r_0 - 1)$.

Alors is est possible de construire pour la suite des systèmes E_r $(r = 2, \ldots, n-1)$ une suite de solutions qui jouirait des propriétés suivantes: 1) chaque système E_r est formé d'équations linéairement indépendantes pourvu que les systèmes précédents soient déja vérifiés; 2) les vecteurs $\vec{x}_{qp}$, $\vec{x}_{n\mu}$ obtenus $(1 \leqslant p \leqslant q < n; \; r_0 \leqslant \mu < n)$ sont linéairement indépendants entre eux.

En vertu du lemme 2, la construction de la suite de solutions peut être terminée par la solution du système (6) pour $r=n$; d'ailleurs cette solution dépend essentiellement de $(n+1)\left[N-\frac{n^3-3n^2}{2(n+1)}\right]$ paramètres arbitraires. Un choix spécial des vecteurs ne joue pas de rôle essentiel, car si ces vecteurs subissent des variations arbitraires mais suffisamment petites, toutes les affirmations restent encore vraies. On obtient ainsi le théorème suivant:.

Théorème. *Chaque espace à n dimensions à connexion affine et sans torsion peut être plongé en un espace affine à N dimensions si*

$$N \geqslant {}^1/_2\,(n^2+2n-1)$$

D'alleurs, en effectuant l'immersion on laisse encore arbitraires $N(n+1)-{}^1/_2\,(n^3-3n^2)$ *fonctions de n arguments.*

En particulier, un espace à deux dimensions et à connexion affine peut toujours être plongé en un espace affine à quatre dimensions.

On a ainsi obtenu un théorème sur l'immersion d'un espace à connexion affine analogue au théorème de Schläefli démontré par Cartan([4]).

En concluant, remarquons que la partie la plus compliquée des raisonnements exposés est la démonstration de l'existence d'une suite régulière de solutions de la suite des systèmes algébriques (6). Dans le mémoire de Cartan ([4]) cette partie des raisonnements est bien simple en vertu du caractère différent du problème considéré. Dans l'article de Shiing-Shen Chern ([5]) cette partie de la démonstration est omise.

Manuscrit reçu
le 8. VIII. 1944.

LITTÉRATURE CITÉE

[1] G. Laptev, C. R. Acad. Sci. URSS, XLI, No. 8 (1943). [2] E. Cartan, Mémoires du séminaire pour l'analyse vectorielle et tensorielle, livr. IV. [3] E. Kähler, Einführung in die Theorie der Systeme von Differentialgleichungen, 1934. [4] E. Cartan, Annales de la société polonaise de mathématiques, 6 (1927). [5] Shiing-Shen Chern, Bull. des Sciences mathématiques, II, 61, 234 (1937).

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

MATHÉMATIQUES

SYSTÈMES DYNAMIQUES SUR UNE MULTIPLICITÉ INTÉGRALE LIMITE

Par V. NIEMYTZKI

(Présenté par A. N. Kolmogoroff, de l'Académie, le 24. XI. 1944)

1. Soit R_t un système dynamique abstrait * défini dans un espace métrique complet R.

Définition 1. Un ensemble $K \in R$, qui ne contient pas de points singuliers de R_t, est nommé multiplicité intégrale limite pour une certaine semi-trajectoire $L \in R_t$ positivement (négativement) stable au sens de Lagrange ** si l'on a: 1) K coïncide avec l'ensemble des points ω (resp. α) limites de L (1); 2) K ne contient pas L. Si les conditions indiquées sont vérifiées, nous dirons que L s'enroule autour de K.

Si R_t est un système dynamique situé sur un plan, on sait, d'après les résultats classiques de Bendixon (2), que K est une courbe simple fermée et tous les mouvements sur K sont périodiques.

Si l'on considère l'espace euclidien à n dimensions pour $n \geqslant 3$, ou bien un espace métrique arbitraire, K peut avoir une structure topologique beaucoup plus compliquée, et les mouvements sur K peuvent être de nature dynamique compliquée.

Le but de la présente Note est d'étudier quelles sont les hypothèses supplémentaires qu'il faut faire sur le mode de convergence de L vers K pour que tous les mouvements sur K soient: 1) récurrents (1), 2) presque périodiques (3).

2. Définition 2. Nous dirons que L s'enroule uniformément sur K si, quel que petit que soit $\varepsilon > 0$, on peut trouver un nombre $T(\varepsilon) > 0$ tel que chaque arc L dont la «longueur temporelle» est au moins égale à $T(\varepsilon)$ est une approximation à ε près de tout l'ensemble K, c'est-à-dire quels que soient les points $p \in L$ et $Q \in K$, on a

$$\rho(f(p; -{}^1/_2 T(\varepsilon),\ +{}^1/_2 T(\varepsilon)),\ Q) \leqslant \varepsilon.$$

Théorème 1. *Pour que toutes les trajectoires contenues dans la multiplicité intégrale K soient récurrentes, il est nécesaire et suffissant que L s'enroule uniformément autour de K.*

---

* On appelle système dynamique abstrait le groupe R_t $(-\infty \leqslant t \leqslant +\infty)$ à un paramètre des transformations de R en lui-même. Le groupe R_t est supposé continu par rapport au point $p \in R$ et par rapport au paramètre t.

** On appelle trajectoire $f(p, t)$ du système R_t l'ensemble des images d'un certain point $p \in R$ pour toutes les transformations du groupe. On appelle point singulier chaque trajectoire qui se réduit à un seul point. Une trajectoire est positivement (négativement) stable au sens de Lagrange si la semi-trajectoire positive (négative) de cette trajectoire est compacte. Une trajectoire est désignée par $f(p, t)$; un arc de trajectoire, considéré comme un ensemble de points, sera désigné par $f(p, t_1, t_2)$.

La condition est suffisante. Supposons que L s'enroule uniformément autour de K. Comme L est stable au sens de Lagrange *, K est compact; donc, en vertu d'un théorème de Birkhoff [1] il suffit de démontrer que K est un ensemble minimal. Pour le voir, il suffit d'établir que chaque semi-trajectoire positive contenue dans K est partout dense sur K.

Supposons, par impossible, qu'il existe un ensemble G ouvert par rapport à K, et une trajectoire $f(x,t)\in K$ telle que l'on a

$$f(x, 0, +\infty)\, G = 0$$

Soit y un point de G; alors, en posant $d=\rho(f(x,0,+\infty), y)$, on a $d\neq 0$. En vertu des conditions du théorème on peut trouver un nombre $T(d/2)$ tel que chaque arc L de longueur temporelle $T(d/2)$ est une approximation de tout l'ensemble K à moins de $d/2$ près. Soit ensuite $\eta>0$ un nombre assez petit pour que la condition $\rho(x,z)\leqslant\eta$ entraîne $\rho(f(x,t), f(z,t))<d/2$ pour $0\leqslant t\leqslant T(d/2)$. Prenons sur L un point R tel que $\rho(x,R)\leqslant\eta$, et prenons sur l'arc $f(R; 0, T(d/2))\subset L$ un point Q tel que $\rho(Q,y)<d/2$; soit $Q=f(R,t_0)$. Soit enfin $Q'=f(x,t_0)$. Nous verrons que le point Q' possède des propriétés contradictoires. En effet, d'une part il suit des conditions $t\leqslant T(d/2)$; $\rho(x,R)\leqslant\eta$ que l'on a $\rho(Q,Q')=\rho(f(R,t_0); f(x,t_0))\leqslant d/2$, et comme $\rho(Q,y)<$ $<d/2$, on a aussi $\rho(y,Q')<d$. D'autre part, $Q'\in f(x,t)$ pour $t>0$, donc $\rho(y,Q')\geqslant d$.

La condition est nécessaire. Supposons que $L=f(p,t)$ ne s'enroule pas uniformément autour de K. Alors il existe un $\varepsilon>0$, une suite d'intervalles à longueur croissante (t_1, t_2), $(t_3, t_4), \ldots, (t_{2n-1}, t_{2n}), \ldots$ et une suite de points $Q_1, Q_2, \ldots, Q_{2n}$ appartenant à K tels que

$$\rho(f(p; t_1, t_2), Q_1>\varepsilon;\ \rho(f(p; t_3, t_4), Q_2>\varepsilon, \ldots, \rho(f(p; t_{2n-1}, t_{2n}) Q_n)>\varepsilon \ldots.$$

Comme K est compact, on peut supposer qu'il existe un point $Q_0\in K$ tel que $\rho(Q_n, Q_0)\to 0$ quand $n\to\infty$, et que $\rho(Q_n, Q_0)<\varepsilon/2$ pour $n=1, 2, \ldots$, c'est-à-dire pour toutes les valeurs de n on a l'inégalité

$$\rho(f(p;\ t_{2n-1},\ t_{2n});\ Q_0)\geqslant\varepsilon/2$$

Considérons la suite des points

$$p_n=f\left(p,\ t_{2n-1}+\frac{t_{2n}-t_{2n-1}}{2}\right)=f(p,\tau_n)$$

comme $t_{2n}-t_{2n-1}\to\infty$ pour $n\to\infty$, on a $\tau_n\to\infty$ pour $n\to\infty$.

Comme L est positivement stable au sens de Lagrange, la suite $\{p_n\}$ possède un point limite p_0 qui est un point ω-limite de L, donc il appartient à K.

Pour simplifier les notations, supposons que $\rho(p_n,\ p_0)\to 0$ pour $n\to\infty$.

En vertu des conditions du théorème chaque trajectoire appartenant à K est récurrente, c'est-à-dire, en vertu d'un théorème de Birkhoff [1], K est un ensemble minimal, donc chaque semi-trajectoire contenue dans K est partout dense sur K. En particulier, $f(p_0,\ \tau)$ pour $\tau>0$ est partout dense sur K. Donc, il existe une valeur τ_0 telle que

$$\rho(f(p_0,\ \tau_0);\ Q_0)\leqslant\varepsilon/4 \qquad (I)$$

Montrons maintenant que cette inégalité contredit à la définition du point p_0. En effet, supposons que $\eta>0$ soit assez petit pour que l'inégalité $\rho(p_0, x)<\eta$ entraîne

$$\rho(f(p_0,\ \tau_0);\ f(x,\ \tau_0))\leqslant\varepsilon/8 \qquad (II)$$

et soit N assez grand pour que l'on ait:

$$1)\ \rho(p_N,\ p_0)<\eta; \qquad 2)\ {}^1/_2(t_{2N}-t_{2N-1})>\tau_0.$$

* Il est facile de voir que pour démontrer la suffisance de notre condition il n'est pas nécessaire de s'appuyer sur la stabilité de L au sens de Lagrange; en effet, la compacité, dont nous avons parlé dans la démonstration, découle immédiatement de ce que L s'enroule uniformément et l'espace R est compact.

Il suit de la seconde condition que le point $f(p_N, \tau_0)$ se trouve sur l'arc $f(p_0; t_{2N}, t_{2N-1})$; donc la distance entre ce point et Q_0 surpasse ε. D'autre part, en vertu du choix de η, il suit des inégalités (I) et (II)

$$\rho(f(p_N, \tau_0), Q_0) \leqslant \rho(f(p_N, \tau_0); f(p_0, \tau_0)) + \\ + \rho(f(p_0, \tau_0), Q_0) \leqslant \frac{\varepsilon}{8} + \frac{\varepsilon}{4} < \frac{\varepsilon}{2}$$

Définition 3. Nous dirons qu'une semi-trajectoire $f(p,t)$; $t>0$ jouit de la propriété S^+ de Franklin si, quel que soit $\varepsilon > 0$, on peut trouver un nombre $\eta(\varepsilon)$ tel que la condition

$$\rho(f(p,t), f(p,t_2)) \leqslant \eta(\varepsilon)$$

entraîne l'inégalité

$$\rho(f(p, t_1+t), f(p, t_2+t)) \leqslant \varepsilon$$

pour chaque valeur de t.

Théorème 2. *Pour que toutes les trajectoires contenues dans la multiplicité intégrale limite K soient presque périodiques, il suffit que la semi-trajectoire L jouïsse de la propriété S^+ de Franklin.*

En effet, soit $\varepsilon > 0$; prenons un nombre $\eta(\varepsilon/2)$ tel que les inégalités

$$\rho(f(p, t_1), f(p, t_2)) \leqslant \eta(\varepsilon/2); \ t_1 > 0, \ t_2 > 0$$

entraînent l'inégalité

$$\rho(f(p, t_1 + t); f(p, t_2 + t)) < \varepsilon/2 \text{ pour } t > 0$$

Soient Q_1 et Q_2 deux points arbitraires sur K tels que

$$\rho(Q_1, Q_2) \leqslant \frac{\eta(\varepsilon/2)}{2} \qquad (I)$$

et soit $t>0$ aussi grand que l'on veut mais fixe. Evaluons la distance $\rho(f(Q_1, \bar{t}), f(Q_2, \bar{t}))$.

Soit $\delta > 0$ assez petit pour que les conditions $\rho(Q_1, x) < \delta$ et $\rho(Q_2, y) < \delta$ entraînent les inégalités

$$\rho(f(Q_1, \bar{t}); f(x, \bar{t})) < \varepsilon/4, \ \rho(f(Q_2, \bar{t}); f(y, \bar{t})) \leqslant \varepsilon/4$$

Supposons d'ailleurs que $\delta \leqslant \frac{1}{4}\eta(\varepsilon/2)$. Prenons sur L des points p_1 et p_2 tels que

$$\rho(p_1, Q_1) < \delta, \ \rho(p_2, Q_2) < \delta \qquad (II)$$

En vertu de (I) et (II) nous avons

$$\rho(p_1, p_2) \leqslant \rho(p_1, Q_1) + \rho(Q_1, Q_2) + \rho(p_2, Q_2) \leqslant \eta(\varepsilon/2)$$

D'après la définition de $\eta(\tau/2)$ et en tenant compte de ce que L jouit de la propriété S^+, nous avons

$$\rho(f(p_1, \bar{t}); f(p_2, \bar{t})) \leqslant \varepsilon/2 \qquad (III)$$

En profitant de ce que δ peut être choisi arbitrairement et en tenant compte de (III), on a

$$\rho(f(Q_1, \bar{t}); f(Q_2, \bar{t})) \leqslant \rho(f(Q_1, \bar{t}); f(p_1, \bar{t})) + \\ + \rho(f(p_1, \bar{t}); f(p_2, \bar{t})) + \rho(f(p_2, \bar{t}); f(Q_2, \bar{t})) \leqslant \varepsilon$$

Comme η ne dépend que de ε, et t est arbitraire, la famille des trajectoires sur K jouit de la propriété S^+, et comme K est compact, chaque trajectoire sur K est stable au sens de Lagrange dans les deux directions; donc, en vertu d'un théorème de Markoff ([3]), elle jouit de la propriété S^-. Donc, toutes les trajectoires sur K jouissent des propriétés S^+ et S^-, et, par conséquent, chaque trajectoire est ou bien périodique, ou bien elle appartient à l'ensemble minimal des mouvements presque périodiques ([3]).

En comparant les théorèmes 1 et 2, on obtient le

Théorème 3. *Pour que toutes les trajectoires contenues dans la multiplicité intégrale limite K appartiennent à une même famille de mouvements presque périodiques, il suffit que* 1) *L jouisse de la propriété S^+, et* 2) *s'enroule uniformément autour de K.*

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Manuscrit reçu
le 24. XI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

MATHEMATICS

ON A PROBLEM OF ANALYTICAL GEOMETRY

By D. K. FADDEJEV

(Communicated by I. M. Vinogradow, Member of the Academy, 27. VII. 1944)

1°. The present note is devoted to the following problem: given in the three-dimensional space a surface of second order

$$f(x_1, x_2, x_3) = \sum a_{ik} x_i x_k = 1$$

with coefficients belonging to a fixed field **R**, to find all the rotations of the space such that the coefficients of the surface remain in **R**. We will suppose that the surface is not one of revolution.

This problem is equivalent to the following one: given a symmetric matrix $\| a_{ik} \|$, $i, k = 1, 2, 3$, with elements belonging to **R**, the lateral roots of which are distinct, to find all the matrices with elements in **R** similar to the given matrix.

Let us introduce the field **k** generated by one of the roots of the characteristic polynomial. To simplify our considerations, we shall suppose that the characteristic polynomial is irreducible, and the field **k** is not cyclic, so that the Galois group of the minimal normal field **K** containing **k** is the complete symmetric group of the third degree. The following considerations, however, are valid in the general case, provided the field **K** is replaced by a suitable algebra of rank 6 over **R**, which is, in general, reducible.

Our purpose is to find all the representations of the field **k** by symmetric matrices, when one of such representations is known. As Gorshkov has shown (1), to any such representation there corresponds a triplet of numbers α_1, α_2, α_3 in **k** orthogonal to conjugate triplets and, conversely, there is always a representation associated with anyone of such triplets. The triplet is determined by the corresponding representation uniquely, up to an arbitrary factor belonging to **k**, and consists of the components of the eigenvector of the matrix representing any primitive element of the field **k**. Thus, in case **k** is a cyclic field, our problem is equivalent to the determination of all the fields with the groups consisting of the quaternion units, that contain the field **k**, as is shown in (2).

Our problem can also be put in the following form. Suppose the field **k** is represented in the space $\mathbf{R}_2$ as a system of points (see (3), chapt. I). Put $\lambda = \alpha_1^2 + \alpha_2^2 + \alpha_3^2$, where α_1, α_2, α_3 is a triplet of numbers orthogonal to the conjugate triplets. We multiply **k** by the point with the coordinates $1/\sqrt{\lambda}$, $1/\sqrt{\lambda'}$, $1/\sqrt{\lambda''}$. Thereby **k** is transformed into a similar system of points $\mathbf{k}_1$ containing the points $\alpha_1/\sqrt{\lambda}$, $\alpha_2/\sqrt{\lambda}$, $\alpha_3/\sqrt{\lambda}$ that are end-points of orthonormalized vectors. In other words, $\mathbf{k}_1$ has cubic symmetry. Conversely, if among the systems of points similar to **k** there is a system $\mathbf{k}_1$ with cubic symmetry, then **k** contains a triplet orthogonal to the conjugate ones. Such a triplet is formed by the numbers which are carried into the end-points of

the mutually orthogonal vectors of equal length by the transformation of $\mathbf{k}_1$ into $\mathbf{k}$ (such vectors necessarily exist in $\mathbf{k}_1$). Thus, our problem may be formulated as follows: to find all the systems of points with cubic symmetry similar to the field $\mathbf{k}$, provided one of such systems is known.

2°. Suppose $U=(\alpha_1, \alpha_2, \alpha_3)$ is the given vector with coordinates belonging to the field $\mathbf{k}$ orthogonal to the conjugate vectors $U'(\alpha_1', \alpha_2', \alpha_3')$ and $U''(\alpha_1'', \alpha_2'', \alpha_3'')$, and $V(\xi_1, \xi_2, \xi_3)$ is any other vector possessing the same property. Resolve the vectors V, V', V'' into their components with respect to U, U', U'':

$$V = \lambda_1 U + \mu_{12} U' + \mu_{13} U''$$
$$V' = \mu_{21} U + \lambda_2 U' + \mu_{23} U''$$
$$V'' = \mu_{31} U + \mu_{32} U' + \lambda_3 U''$$

Here $\lambda_1 = \alpha_1\xi_1 + \alpha_2\xi_2 + \alpha_3\xi_3 \in \mathbf{k}$, $\lambda_2 = \lambda_1'$ and $\lambda_3 = \lambda_1''$ are conjugate to λ_1, $\mu_{12} = \xi_1\alpha_1' + \xi_2\alpha_2' + \xi_3\alpha_3'$ belongs to the normal field $\mathbf{K}$, while the other μ_{ij} are conjugate to μ_{12} by the corresponding substitutions of the Galois group. In virtue of orthogonality of V, V', V''

$$\begin{aligned} \lambda_1\mu_{21} + \lambda_2\mu_{12} &= -\mu_{13}\mu_{23} \\ \lambda_1\mu_{31} + \lambda_3\mu_{13} &= -\mu_{12}\mu_{32} \\ \lambda_2\mu_{32} + \lambda_3\mu_{23} &= -\mu_{21}\mu_{31} \end{aligned} \qquad (*)$$

Conversely, having chosen any number $\mu_{12} \in \mathbf{K}$ under the only condition $\mu_{21}\mu_{13}\mu_{32} + \mu_{31}\mu_{12}\mu_{23} \neq 0$, we obtain λ_1, λ_2, λ_3 from (*) such that λ_1 belongs to the field $\mathbf{k}$, and λ_2 and λ_3 are conjugate to λ_1. Putting $V = \lambda_1 U + \mu_{12} U' + \mu_{13} U''$, we obtain a vector orthogonal to the conjugate vectors, the coordinates of which belong to $\mathbf{k}$. It is not difficult to see that, in case the field $\mathbf{R}$ is real, the system (*) is inconsistent, as soon as its determinant $\mu_{21}\mu_{13}\mu_{32} + \mu_{31}\mu_{12}\mu_{23}$ is 0. Consequently, the method just exposed enables us to find all the vectors (ξ_1, ξ_2, ξ_3) that are orthogonal to the conjugate ones, as soon as one of such vectors $(\alpha_1, \alpha_2, \alpha_3)$ is known.

3°. Let us prove the following

Theorem. *In order that a system of points, which represent a cubic field* $\mathbf{k}$, *have cubic symmetry, it is necessary and sufficient that the discriminant of the field* $\mathbf{k}$ *be represented by the quadratic form* $x^2 + 2y^2$.

Proof. Suppose ρ is a number from the field $\mathbf{k}$, whose spoor is 0, and $\rho^3 = q\rho + n$ is the equation, a root of which is ρ. Consider the numbers $\lambda = \frac{3\rho^2 - 2q}{q}$, $\mu = \frac{3n\lambda - 2q\rho}{j}$, where j is the index of the number ρ. Then $\mathrm{Sp}\,\lambda = \mathrm{Sp}\,\mu = \mathrm{Sp}\,\lambda\mu = 0$, $\mathrm{Sp}\,\lambda^2 = 6$, $\mathrm{Sp}\,\mu^2 = \frac{2(4q^3 - 27n^2)}{j^2} = 2d$, where d is the discriminant of the field $\mathbf{k}$. The numbers 1, λ and μ can be considered as a basis of the field $\mathbf{k}$. We have $\mathrm{Sp}\,(t_1 + t_2\lambda + t_3\mu)^2 = 3t_1^2 + 6t_2^2 + 2dt_3^2$. Hence, in order that the field $\mathbf{k}$ possess cubic symmetry, it is necessary and sufficient that the quadratic form $d(3t_1^2 + 6t_2^2 + 2d\,t_3^2)$ be contained in the quadratic form $u_1^2 + u_2^2 + u_3^2$. The last condition is fulfilled if and only if $3d$ is a norm in the field $\mathbf{R}(\sqrt{-2})$, *i. e.* d is represented by the quadratic form $x^2 + 2y^2$, as was to be proved.

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Received
27. VII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

AERODYNAMICS

ON SOME FUNDAMENTAL RELATIONS IN AERODYNAMICS OF THE TURBULENT BOUNDARY LAYER IN COMPRESSIBLE AND INCOMPRESSIBLE FLUIDS

By A. M. FEINSILBER

(Communicated by A. N. Kolmogoroff, Member of the Academy, 18. VIII. 1944)

The linear law of distribution of friction, which is valid within the scope of the inner problem of the theory of turbulence (flow in circular tubes), cannot be applied to a turbulent boundary layer. For the latter a different law of distribution of friction must therefore be established. The purpose of the present note is to complete the theory of the turbulent layer, and it is by directly satisfying the differential equation of the layer that the missing relation is supplied.

We adopt the following notations: u for the longitudinal velocity component in the layer, ρ for the density, p for the pressure, τ for the friction stress in the layer, τ_0 for the friction stress at the wall, $\bar{u}$ and $\bar{\rho}$ for the respective values of the velocity and the density at the boundary of the layer, δ for the thickness of the layer.

For the independent variables we take x and $t=y/\delta$, and introduce the dimensionless functions

$$u_1=u/\bar{u}; \quad \rho_1=\rho/\bar{\rho}; \quad \tau_1=\tau/\tau_0$$

Let us first consider the turbulent boundary layer in a compressible fluid on a plate. In this case we may regard u_1 and ρ_1 as independent of x, *i. e.* as functions of t only.

The required relation is found in the form

$$\tau_1=\frac{\tau}{\tau_0}=1-\frac{\int_0^t \frac{du_1}{dt}\int_0^t \rho_1 u_1\,dt\,dt}{\int_0^1 \frac{du_1}{dt}\int_0^t \rho_1 u_1\,dt\,dt} \qquad (1)$$

It is obtained by eliminating the quantity $d\delta/dx$ from the relations equivalent to the differential equation of the boundary layer and the theorem of momentum. What is essential is that (1) involves neither the thickness δ of the boundary layer, nor, and this is particularly important, its derivative $d\delta/dx$ (this last quantity is experimentally not determinable at all). Thus, the determination of the distribution of velocities and friction and the determination of the thickness of the layer are now separate problems.

In the particular case of an incompressible fluid we have to put $\rho_1 = 1$, and formula (1) may be rewritten in the form:

$$\tau_1 = \frac{\tau}{\tau_0} = 1 - \frac{\int_0^t \frac{du_1}{dt} \int_0^t u_1 \, dt \, dt}{\int_0^1 \frac{du_1}{dt} \int_0^t u_1 \, dt \, dt} \tag{2}$$

By comparing formula (1) and, correspondingly, (2) with the laws already known we may find the distribution of velocities and friction. The following method, for example, may be suggested to determine the distribution of friction stresses in the turbulent layer in an incompressible fluid. Assume the usual non-universal law $u_1 = t^n$ [1], where n is a function of Reynolds' number Re varying from $n = \frac{1}{7}$ for $\mathrm{Re} = 2 \cdot 10^4$ to $n = \frac{1}{10}$ for $\mathrm{Re} = 2 \cdot 10^6$. Then by formula (2) we obtain the law of friction distribution

$$\tau_2 = \sqrt{\tau_1} = \sqrt{1 - t^{2n+1}} \tag{3}$$

differing considerably from the corresponding law for the tube.

It is noteworthy that the law (3) is practically universal (*i. e.* in the segment $0 \leqslant t \leqslant 1$, corresponding to the boundary layer, τ_2 is practically on stant when n varies from $\frac{1}{7}$ to $\frac{1}{10}$).

The comparison of formula (3) with Prandtl-Karman's law [2] of the mixing path yielded the following distribution of velocities ($\varkappa$ being an experimental constant):

$$\frac{\bar{u} - u}{\sqrt{\tau_0/\rho}} = \frac{0.507}{\varkappa} \left[\ln \frac{1 + 0.0418\,[(1-t) + \sqrt{1-t}]}{(1 - \sqrt{1-t})^2} - \right.$$
$$\left. - 10.0 \operatorname{arctg} \frac{0.5 + \sqrt{1-t}}{4.87} + 1.05 \right] \tag{4}$$

For an arbitrary profile the relation analogous to (1) naturally assumes a considerably more complicated form. If, for instance, the velocity distribution for an incompressible fluid is again taken to be $u_1 = t^n$, we obtain the following law of friction distribution in the layer:

$$\tau_1 = (1-t) + \left(1 + \frac{\delta \frac{dp}{dx}}{\tau_0}\right) t\,(1 - t^{2n}) \tag{5}$$

The exponent n may be considered as a new experimental constant. Its value found by comparison with experimental data is $n = \frac{1}{2}$.

Once the law of velocities is established, the resistance law can be found immediately.

Received
8. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

MATHEMATICAL PHYSICS

CONCERNING THE JUSTIFICATION OF REPLACEMENT OF TIME AVERAGES BY PHASE AVERAGES IN STATISTICAL MECHANICS

By J. P. TERLEZKI

(Communicated by A. N. Kolmogoroff, Member of the Academy, 7. VI. 1944)

It is usually considered that in order to justify the method of phase averages (Gibbs' method) in theoretical computations of averages of physical quantities observed experimentally, it is necessary to prove the equality of time averages and phase averages, *i. e.* the ergodene hypothesis. It is, however, known that this last proposition not only cannot be proved under general assumptions, but is actually false for a large class of dynamical systems, for instance, for a system performing a conditionally periodic motion, a particular case of which is the linear system of connected oscillators. However, the fruitfulness of Gibbs' method in application to various physical systems calls for further attempts to prove its lawfulness.

A simple physical analysis of the phenomena and experiments to which statistical mechanics is applicable shows that in proving the method of phase averages there is no need to require that time averages be exactly equal to phase averages. It will suffice, for instance, to require an approximate equality between the averages for a large majority of the initial values which the dynamical variables (the coordinates and impulses of all material points) can possibly take under the conditions of the experiment. What, moreover, can and must be used in the proof, is that all systems to which methods of statistical mechanics are applied, possess a large number of degrees of freedom.

The fact that in phase domain of admissible initial values of very small measure there is a considerable divergence between time and phase averages presents no obstacle to employing the method of phase averages for the computation of experimentally observed quantities. In fact, in all cases of application of statistical mechanics to real observable systems, these last are separated from the outer world, *i. e.* from a certain large system with practically an infinite number of degrees of freedom, by fixing a finite number of macroscopical parameters. Thus, the very arbitrary choice of initial conditions is inherent in the experiment, and this makes acceptable any theory which is true for a large majority of macroscopically identical experiments.

As a sufficient condition for the applicability of the method of phase averages to the computation of the average value of $F(X)$, where X are the canonic variables of an isolated system, we may take proceeding from the above considerations that *

$$\overline{(\overline{F^t} - \overline{F})^2} \to 0 \text{ for } N \to \infty \qquad (1)$$

where N is the number of the degrees of freedom,

* A similar condition was considered in connexion with analogous questions by A. Khintchine (1).

$$\overline{F}^t(X_0)=\lim_{T\to\infty}\frac{1}{T}\int_0^T F(X_0,t)\,dt \quad \text{(the time average)},$$

$$\overline{F}=\int W(X)F(X)\,dX=\int W(X_0)F(X_0,0)\,dX_0 \quad \text{(the phase average)}.$$

X_0 is the initial value of X for $t=0$; $W(X)$ is the stationary function of the probability density; it is determined by means of the separation of the system from the external world.

Consider an isolated linear system N of oscillators connected to each other. We shall show that for the most interesting physical quantities (the kinetic and potential energy, the products of coordinates and impulses, *etc.*), which are quadratic functions of the coordinates and the impulses of a finite number of oscillators the condition (1) is satisfied. The only condition, beside the physically natural ones, which we shall impose in the proof on $W(X)$ is that the independent motion integrals of the system will be assumed to be pairwise statistically independent. This condition is satisfied, among other possible probability distributions, by both the canonic and microcanonic distribution for $N\to\infty$.

Let us denote by x_k the coordinates and the impulses of the oscillators under consideration. Let y_n be the normal coordinates of the whole system, and ω_n, the corresponding cyclic frequencies. Then

$$x_k=\sum_{n=1}^{N} a_n^k y_n, \quad \text{where} \quad y_n=B_n\cos(\omega_n t+\varphi_n) \tag{2}$$

the a_n^k being the constant coefficients determined by the Hamiltonian of the system; B_n and φ_n, the amplitudes and the phases of normal oscillations.

Passing to cyclic coordinates — the action variables $I_n=\frac{1}{2}B_n^2\omega_n^3$ and angle variable $w_n=\omega_n t+\varphi_n$, we observe that the time averages of the considered functions of x_n are equal to the phase averages, but only with respect to the angle variables w_n, since in these coordinates the stationary function does not depend on w_n and is a function only of the I_n, *i. e.* $W(I_1, I_2, \ldots, I_N)$.

Consider the time and phase averages of the quantity $F(X)=x_k(t)\,x_l(t+\tau)$, *i. e.* of the product of x_k and x_l in different moments of time. The physically interesting quantities mentioned above are sums of this kind of products for $\tau=0$.

After simple calculations we find for the time average of this quantity

$$\overline{F}^t=\overline{x_k(t)\,x_l(t+\tau)}^t=\sum_{n=1}^{N} I_n\frac{a_n^k a_n^l}{\omega_n^2}\cos\omega_n\tau \tag{3}$$

whence for the phase average we obtain

$$\overline{F}=\overline{x_k(t)\,x_l(t+\tau)}=\sum_{n=1}^{N}\overline{I}_n\frac{a_n^k a_n^l}{\omega_n^2}\cos\omega_n\tau \tag{4}$$

Inasmuch as the I_n are independent motion integrals, we may in view of the assumption made with respect to $W(X)$, consider the right-hand side (3) as a sum of pairwise independent random variables, and the right-hand side of (4) as the sum of its mathematical expectations. But for such sums (under the natural assumption that they remain finite for $N\to\infty$) the law of large

numbers is valid. Consequently, the probability of the inequality $|\overline{F^t} - \overline{F}| > \varepsilon$ vanishes for arbitrarily small ε, if $N \to \infty$. The condition (1) is thus also satisfied, and our assertion is proved.

This theorem may be easily generalized to the case of an arbitrary system performing a conditionally periodic motion. Just as in the preceding case, we may prove by passing to action variables and angle variables that for a wide class of functions $F(X)$

$$\overline{F} = \int W(I)\,\overline{F^t}(I)\,dI \tag{5}$$

where I is the whole aggregate of the action variables. Hence to all quantities satisfying the condition

$$\overline{F^t}(I) = \frac{1}{N}\sum_{i=1}^{N} \varphi_i(I_i) \tag{6}$$

where the $\varphi_i(I_i)$ are functions of I_i, the law of large numbers may be applied, if the I_i are assumed to be pairwise statistically independent. Thus, for quantities of the type (6) the condition (1) is satisfied.

The concrete examples discussed above show that the condition (1) which is physically acceptable in proving the method of phase averages, is satisfied in those real physical cases, to which Gibbs' method is applied, but in which the ergodene hypothesis in its usual form is manifestly false.

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Received
7. VI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PHYSICS

ON THE HARDENING OF METALS

By G. V. AKIMOV, Corresponding Member of the Academy

1. It is a known fact that pure single crystal metals are characterized by extremely low strength and extremely high plasticity. Thus, single crystal of pure aluminium has a hardness (H_{Br}), ultimate shearing strength (τ_B) and residual shear at fracture (φ) 15—20 kg/mm², 4—5 kg/mm² and 110 per cent, respectively, while the corresponding figures for single crystal of pure zinc are: 20—40 kg/mm², 1.2 kg/mm² and 350—500 per cent.

It is also a well known fact that metals may be hardened, and that commercial metals, combining high strength with a certain reserve of plasticity, are to one degree or another hardened metals. Hardening of metals is achieved by the following methods:

A. Transition from pure single crystal metals to polycrystalline metals containing varying quantities of metallic and non-metallic impurities.

B. The formation of solid solutions.

C. The formation of binary structures of the type: soft basic phase of nearly pure metal or solid solution—solid inclusions (generally intermetallic compounds).

D. The origination of metastable structures of the type of martensite.

E. Deformation.

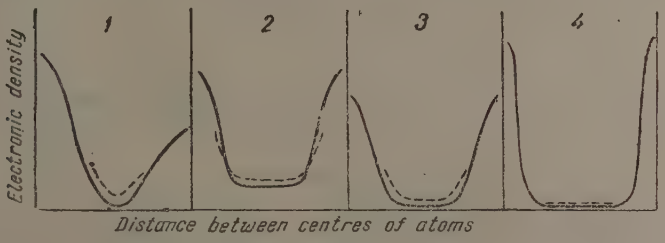
2. To explain the nature of hardening, several hypotheses have been put forward. Here mention may be made of the hypothesis of pins and interference gliding ([1]), the numerous versions of the hypothesis of «gearing» and distortion of the lattice ([2-4]), the old hypotheses that a brittle amorphous layer may appear on grain boundaries and on glide planes ([5,6]), and the hypothesis that the atoms themselves may change during deformation ([7]). Of late much attention has been given to the «dislocation» hypothesis ([8]), despite its complicated and artificial nature. Frequently, especially in researches concerned with the physics of metals, hardening is connected with the presence of internal stresses in the lattice of the metal (so-called stresses of the third order).

3. The development of the theory on chemical bonds in crystals, based on the views of Bragg ([9, 10]), who proposed that a crystal be regarded as a three-dimensional Fourier series, makes it possible, in our opinion, to interpret hardening in metals from the general physical-chemical point of view.

Bragg's views, in conjunction with the conceptions of quantum mechanics on the structure of atoms, make it possible to regard the interparticle bond as the presence of a certain density of electrons between the centres of the particles. If the valency electrons are rigidly bound to definite points in the crystal, the bond is of a directed, orientated nature. To this category belong all types of bond, except metallic. In the case of the latter the valency electrons are to a considerable degree free and the bond has no direction (rigidity). In general the type of

bond is roughly determined by the nature of the change in density of the electrons between the centres of the particles or, in other words, by the nature of the distribution of the electron density in the crystal and by the fixedness or, inversely, freedom of valency electrons.

If a continuous rigid «skeleton» with a high electron density (of the order of single electrons per cubic angstrom) exists within the crystal, the bond is called a t o m i c. A characteristic example of a bond of this type is the diamond. If the regions of high electron density in the crystal alternate with regions of very low density (of the order of hundredths and tenths of an El/Å^3) and, hence, there is no rigid «skeleton», the bond is of the ionic or molecular type. The latter type



Distribution of the electron density in crystals with different types of bond. The dotted curve represents diagramatically the highest values of the density, and the solid curve, the lowest. *1*, ionic bond; *2*, atomic bond; *3*, molecular bond; *4*, metallic bond

is characteristic for many organic compounds; the NaCl crystal is a well-known example of the ionic bond.

The metallic bond, besides the fact that it does not have direction (there is the possibility of the electrons moving freely about), is characterized by the nearly uniform filling of the interatomic space by an «electron cloud»; here the electron density has low values (of the order of tenths and hunhredths of an El/Å^3).

There is also the possibility of the existence of a mixed bond, when a bond of one type exists along one direction in the crystal, and along the other directions the linkage is of a different type. A characteristic example of a mixed bond is graphite, in which an atomic bond acts along two coordinates, and a metallic bond, along the third. The figure shows diagramatically the change in the electron density between particle centres for the basic types of bond.

Bragg's views have made it possible not only to find the proper quantitative expression for the conception of the interparticle bond, but also to discover an experimental method with which one may quantitatively determine the electron density, and, hence, get at the nature of the bonds in crystals. With this method the electron distribution was determined in crystals of NaCl, diamond, hexamethylene-tetramine, magnesium ([10]), aluminium and several other substances.

4. The unique physical-chemical properties of metals are connected with the specific character of the metallic bond. High electrical conductivity, as well as high plasticity is due to the presence of free electrons, through which linkage of the metal ions in the solid or liquid state is effected (M^+ — electron — M^+...) It is possible to considerably deform a metal in the solid state and displace certain parts of its lattice with respect to others without disturbing the bonds between the ions, since the mobile, free electrons, penetrating the entire lattice, maintain linkage between the metal ions along all directions all through the deformation process. Only after appreciable deformation (and hardening that sets in) does the metal's reserve of plasticity become exhausted and then its lattice is no longer safe from brittle destruction. In general, the closer the lattice of a metal is to the ideal metallic lattice, the more pronounced do the typical properties of the metallic bond become.

5. The proposed hypothesis for the hardening of metals consists of the following statements.

A. The absorption of energy by the metal upon deformation (plastic deformation by outside forces, internal deformation caused by penetration of the solid solution by foreign atoms, the falling out of particles from the solid solution, *etc.*) or near some impurities (of a chemical or physical nature) in the metal lattice (inclusions, boundaries of the bloks and individual grains, impurities in the lattice, imperfections in the lattice and other deviations from an «ideal» metal) can transform a metallic bond to an atomic one.

The nearly uniform distribution of the free electrons («electron gas») in the regions where transformation of the bond occurred, is replaced by a rigid «electron bridge». This is accompanied by an increase in electron density in these regions and a corresponding decrease in others, while the electrons themselves are found to be rigidly bound with a definite region. It is possible that the effect ressembles the formation of plane and linear double layers at the interfaces within the metal.

The atomic bond corresponds to a higher energy level in comparison with the metallic bond. During plastic deformation, the metal absorbs a considerable quantity of energy, of the order of 1 Cal/gram at maximum degrees of deformation. Only a small part of this energy is contained in the metal in the elastic form, a large part of it is in some unknown form. According to our assumption, bond transformation is also accompanied by a change in the energy level. Thus, the metal may absorb energy while rearranging its bonds.

B. The strengthening of metal is explained by the transformation of the metallic bond into an atomic one, *i. e.*, by the formation of a rigid continuous electron skeleton in a medium consisting of a nearly uniform «electron cloud».

In a very plastic metal under strain when interfaces of diverse nature arise and absorption of energy takes place, one may observe the appearance of regions with transformed linkage whose properties markedly differ from those of a typical metal. These regions possess high strength and hardness and a low plasticity, *i. e.* ressemble the typical exponents of an atomic bond—the diamond and corundum. If the regions with an atomic bond that have arisen in a given volume of the metal are joined into a continuous carcass, the volumes which still have a metallic bond are completely blocked and the metal behaves practically like a rigid brittle solid. If, on the other hand, the regions with transformed linkage are scattered through the metal without forming a continuous carcass, then during deformation (and only during deformation do we determine the strength and plasticity of a solid) the metal behaves like a body that is still plastic, capable of a certain amount of further plastic deformation. Only when in the course of deformation at the expense of further bond transformation a skeleton with altered bonds arises, that passes through the entire volume (which corresponds to exhaustion of plasticity), does the metal begin to behave like an elastic solid and breaks down (at a certain «peak») when the rupture stress attains a value corresponding to the limiting theoretical strength. Intermetallic compounds (transition phases of the type of carbides, nitrides, *etc.*) seem to come very nearly under the scope of the first case, *i. e.* the formation of a continuous rigid skeleton even before deformation. Hardening at the expense of the formation of binary structures corresponds most frequently to the second case, *i. e.* to the absence of a continuous skeleton; here hardening appears only in the process of deformation. Metastable (including martensite) structures, depending upon the conditions of their formation, may be close to both the first and the second type. Of course, the magnitude of the transformed regions and their distribution in the metal will play a very important rôle. It may be particularly pointed out that, according to our hypothesis, the hardening is rather connected with the changes in the existing electron medium of metal than with breaking of more profound electron layers of the metallic ions.

Upon heating, *i. e.* upon increasing the kinetic energy of the particles, the atomic bond may be retransformed into the initial metallic one. Crystals of Fe_3C and many other carbides decompose in the process of sufficiently high and pro-

tracted heating into the metal and graphite. Martensite upon tempering is transformed into binary structures corresponding to a smaller degree of transformation of the metallic bonds. Metals that have been hardened by deformation during heating, disclose known phenomena of relaxation and recrystallization, which, in accordance with our hypothesis, also represent restoration of the metallic bond.

Indirect quantitative confirmation of the hypothesis is not only supplied by the change in the mechanical qualities (lowering in plasticity at hardening), but also by the change in certain other properties, namely, the rise in the specific electrical resistance and coercive force and also a lowering in the electric potential by hardening.

However, a direct check of the hypothesis can probably only be made on the basis of new X-ray technique (precise measurements of the X-ray intensities), as well as on the basis of a detailed optical and magnetic investigation of soft and hardened metals.

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Received
15. XI. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PHYSICS

ON THE DEPOLARIZATION OF LIGHT SCATTERED IN SOLUTIONS OF RUBBER-LIKE POLYMERS

By V. TSVETKOFF and E. FRISMAN

(Communicated by S. I. Wawilow, Member of the Academy, 31. VII. 1944)

Investigation of the depolarization of light, Δ, scattered by colloid solutions is one of the methods of determining the dimensions, shape and optical properties of the particles suspended in them.

A study of the state of polarization of the scattered light is found to be especially helpful in this respect. If the incident beam is directed along the x-axis of a rectangular system of coordinates and observation is carried out along the y-axis, the degree of depolarization of the scattered light is usually determined by the ratio $\Delta = I_x/I_z$ where I_x and I_z are the x- and z-components, respectively, of the scattered light. Here one should discriminate between the quantity Δ_v (the electrical vector of the incident light parallel to the z-axis) and the quantity Δ_h (the electrical vector of the incident light parallel to the y-axis).

On the basis of theories advanced by Rayleigh (1), Mie (2) and Gans (3) one may draw the following general conclusions regarding the relation between Δ and the properties of the dispersed phase.

For spherical optically isotropic particles, irrespective of their dimensions, $\Delta_v = 0$. If the particles are anisotropic in shape or in their optical properties, $1 > \Delta_v > 0$.

In a system which does not contain large particles (*e. g.*, for all molecular liquids) $\Delta_h = 1$, irrespective of their anisotropy and shape. On the other hand, when the dispersoid is comparable in dimensions with the wave-length of the incident light, $\Delta_h > 1$.

It was found by Krishnan (4) that $\Delta_h > 1$ for binary liquid mixtures as well as colloid systems, which he interpreted as proof of the existence in the former of large molecular swarms (the Krishnan effect). To explain this effect Gans (5) proposed a theory in which he assumed that the fluctuations in the density and anisotropy of the various volume elements are correlated, and that the linear dimensions of a swarm are small in comparison with the wave-length of the light. Gans arrives at the conclusion that the Krishnan effect ($\Delta_h \neq 1$) may exist only for the case of a non-spherical swarm, whereas for spherical particles, regardless of their optical anisotropy, $\Delta_h = 1$. In the particular case of rod-like particles of length l

$$\Delta_h = 1 - \frac{4\pi^2}{45}\left(\frac{l}{\lambda}\right)^2 \left[1 - 3\,\frac{(g_1 - g_2)^2}{(g_1 - g_2)^2 + (g_2 - g_3)^2 + (g_1 - g_3)^2} \right] \qquad (1)$$

where g_1, g_2 and g_3 are quantities proportional to the difference between the dielectric constants of the medium and the particle along its three mutually perpendicular axes (g_3 in the direction of the long axis of a rod).

Upon making the plausible assumption for a rod-like particle that $g_1 = g_2 \neq g_3$, Gans obtained

$$\Delta_h = 1 - \frac{4\pi^2}{42}\left(\frac{l}{\lambda}\right)^2 < 1 \qquad (2)$$

However, since for all systems that had been investigated until then experiment gave $\Delta_h \geqslant 1$, Vrkljan and Katalinič ([6]) corrected Gans' results and, assuming that a thread-like particle possesses highly improbable optical properties ($g_1 \neq g_2 = g_3$), obtained with the aid of (1) $\Delta_h > 1$. Gehman and Field ([7]) were apparently the first to observe the scattering in which $\Delta_h < 1$, while studying solutions of natural rubber.

Since in natural rubber a considerable quantity of colloidal admixtures is always present which may greatly distort the experimental results, we employed solutions of some synthetic polymers to study the depolarization of the scattered light, paying attention to the degree of purity of the solution. The following polymers were investigated: technical natural rubber (smoked sheet), polystyrol, divinyl polymers, and isobutylene polymers (oppanol and vystanex).

Polystyrol was obtained by polymerizing monostyrol that had been distilled in a closed vessel and did not contain any noticeable quantity of scattering admixtures. The solutions of polystyrol, as well as those of oppanol and vystanex, were purified simply by letting them stand for several days so that any admixture may settle. The absence of any noticeable quantity of colloidal admixture was established by a control measurement of their electrical double refraction. The Kerr effect in the field $E = 40\,000$ volts/cm did not noticeably exceed that for the pure solvent (CCl_4).

The divinyl rubbers were fractions (of different molecular weight μ) of one and the same polymer, specially prepared and containing the least possible amount of impurities. Their solutions were purified by centrifuging. Natural rubber was found to contain the most colloidal admixtures. Its solution could not be purified by ordinary centrifuging, and we had to resort to a method that we employed elsewhere ([8])—multiple action of a strong electrical field with subsequent centrifuging. This enabled us to prepare clear solutions containing a highly insignificant amount of foreign particles. The presence of the latter was established from the birefrigence of the solution in an electric field. Table 1 shows the observed values of birefringence Δn in an electric field E for a solution of natural rubber in CCl_4. We see that after the solution is acted upon by the electric field for three hours and then centrifuged, the electro-optical effect decreases tenfold, although it is still several times the value of the Kerr effect for pure tetrachlormethane.

Table 1

Dependence of the Electric Birefringence of a Solution of Rubber in Tetrachlormethane upon the Degree of Purification

Unpurified solution		Solution after purification by applying a field of 40 000 volts/cm for a period of 3 hrs and centrifuging for 2 hrs	
E (in volts/cm)	$\Delta n \times 10^8$	E (in volts/cm)	$\Delta n \times 10^8$
3600	1.8	9000	~0
6600	1.8	18000	0.09
9600	1.7	25500	0.24
14100	2.7	30000	0.33
18900	3.6	34200	0.49
37500	6.5	39000	0.68

The degree of depolarization of the scattered light was determined in the usual way, employing Wollastone's prism and an analyser on the limb (⁹). The results obtained are presented in Table 2. As the data show, the solutions of all the polymers are characterized by an insignificant value of Δ_v, and an anomalous depolarization $\Delta_h < 1$.

Table 2

Degree of Depolarization of Light Scattered in Solutions of Various Polymers

Polymer, solvent, concentration (by volume)	Degree of purification	Δ_v	Δ_h
Natural rubber — tetrachlormethane, $c_v = 0.8\%$	Unpurified	0.020	0.51
	Purified by applying a field for 3 hrs and centrifuging for 2 hrs	0.015	0.27
Divinyl synthetic rubber — ether, $c_v = 2\%$			
Fraction 1; $\mu = 22000$	Before centrifuging	0.032	1.22
	After centrifuging	0.025	0.33
Fraction 2; $\mu = 40000$	Before centrifuging	0.028	0.93
	After centrifuging	0.022	0.31
Fraction 3; $\mu = 61000$	After centrifuging	0.049	0.49
Polystyrol — dichlorethane, $c_v = 2.5\%$	After settling for several days	0.037	0.69
Polystyrol — tetrachlormethane, $c_v = 2.5\%$	Do	0.016	0.40
Vystanex - pentane, $c_v = 0.8\%$; $\mu = 55000$	»	0.022	0.30
Oppanol—pentane, $c_v = 0.8\%$; $\mu = 150000$	»	0.015	0.17

The degree of depolarization of the scattered light depends upon the purity of the solutions. The value of Δ_v decreases somewhat with purification, but the presence of foreign (coarse) particles in the solution especially affects the value of Δ_h, the purer is the solution, the greater Δ_h differs from unity. In a number of cases, where the solution contained a great number of colloidal admixtures, we even obtained $\Delta_h < 1$ (*e. g.* fraction 1 of the divinyl polymer before centrifuging).

Thus, depolarization $\Delta_h < 1$ cannot be regarded as a chance phenomenon due to accidental causes, but is evidently a property that is general for solutions of all rubber-like polymers, the more pronounced, the purer the solution with respect to coarse admixtures.

It is not surprising that such a state of polarization had not been observed in any colloidal system, investigated heretofore, for the properties that the scattering particles should possess according to Gans' theory—a considerable elongation in shape, dimensions smaller than the wave-length of the light, and negligible optical anisotropy—are the characteristic properties of chain-like molecules in solution (¹⁰), and are much more seldom met with together in other colloidal systems.

It is our opinion that the phenomenon considered by Gans in his theory and experimentally discovered by Gehman and Field and by us, differs in principle from the Krishnan effect, since the basic condition for the origination of the latter ($\Delta_h < 1$) is coarse dispersity of the dispersion medium (the dimensions of the particles are comparable with the wave-length of the light), while the effect described in the present paper is connected with the asymmetry of the particle shape. Therefore it should be properly named the Gans effect.

A study of the Gans effect in solutions of high molecular compounds is of undoubted interest, since it promises to furnish direct data on the length of the molecular chain in them. However, the results presented in Table 2 show that in a quantitative interpretation of the experimental data one should be cautious, paying serious attention to the purity of the solution investigated.

Here we shall apply the theory only to the solutions of polyisobutylenes, since they are the purest with respect to colloidal admixtures, and the Gans effect for them is particularly great. The latter circumstance is quite characteristic, since polyisobutylenes differ from other polymers by the absence of branching in the molecule, which is quite elongated, very much like a rod.

The length of a scattering particle in the solution may be determined with the aid of equation (2). Inserting $\Delta_h = 0.30$ for vystanex and $\Delta_h = 0.17$ for oppanol, we get $l = 4.2 \cdot 10^{-5}$ and $l = 4.6 \cdot 10^{-5}$, respectively. The results obtained for the dimensions of a scattering cell are too large for one molecule. It is possible that in stationary solution after settling the molecular chains interlace forming micells which are considerably lower than an individual modecule.

The solutions of all the polymers are characterized by a small degree of depolarization Δ_0 (from 1 to 5 per cent). This signifies that the optical anisotropy of their molecules in the solution is negligible. This conclusion is in good agreement with the results of an investigation of the birefrigence of solutions while in flow([10]) and in a magnetic field ([12]), and also with the theoretical computation of Kuhn ([11]).

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

TECHNICAL PHYSICS

THE RADIOLUMINESCENCE OF POTASSIUM URANYL SULPHATE AS THE BASIS FOR A LOW BRIGHTNESS STANDARD

By T. V. TIMOFEYEVA

(Communicated by S. I. Wawilow, Member of the Academy, 11. IX. 1944)

1. That crystals of potassium uranyl sulphate (hereafter referred to for brevity as PUS) luminesce under the action of α-particles emitted in the radioactive disintegration of uranium and its products, was first detected by Becquerel (1). This luminescence was employed by Rayleigh (2) (the younger) as a standard in studying the luminescence of the night sky. Though its brightness is low (of the order of 10^{-4} apostilbs), the luminescence is constant owing to the fact that the chemical change under natural conditions is insignificant, and disintegration of uranium is very slow.

Of late, especially in connexion with the demands of blackout technique, the measurement of low brightnesses and the need for an appropriate standard became quite urgent. This prompted us to undertake an experimental investigation of the luminescence of PUS under the action of its own radioactive radiation.

2. The very first experiments revealed that the brightness of the autoluminescence of two specimens we had at our disposal was quite different, whereas the brightness of their photoluminescence under the action of an external light source was the same.

One specimen (Kahlbaum), in the shape of a plate formed by crystals that had grown together, was some 30 times brighter than another (Central Research Institute for Geological Prospecting) which was in the form of a powder. The radiation of α-particles in both specimens was exactly the same and was governed entirely by the α-disintegration of the uranium itself, since the radium content, according to a determination made in the geochemical laboratory of the Radium Institute of the Academy of Sciences, was found to be of the same order in both specimens, namely 10^{-12} g of radium per gram of uranium, which amounts to some hundredths of a per cent of the α-radiation due to uranium.

Neither can the difference in brightness be explained by a difference in layer thickness and in the weight of the sample examined, since luminescence still was weaker in the case of the powder even when an amount equal in weight and thickness to the plate was taken. Further experiments showed that it is due to the difference in dispersion between the two specimens and, accordingly, the greater scattering of light in the powder than in the crystals forming the plate, for upon pulverizing the latter weak autoluminescence was exhibited by both.

3. To check this conclusion we carried out some immersion experiments. The powder specimen was immersed in liquids that do not dissolve it and whose refractive indexes are very close to its own. This increased brightness by 30 to 90 per cent depending upon the liquid, namely, 30—40 per cent for ethyl alcohol, and 70—90 per cent for benzene.

Since the reftractive index of benzene is practically equal to that of the uranyl salt, this is probably the maximum increase in brightness that can be achieved by immersion.

Immersion in a liquid leads, in the first place, to a lowering in the scattering of the light due to radioluminescence, and hence, to a decrease in its absorption by the powder, since the emission and absorption spectra overlap. In the second place, immersion favours an increase in the visible brightness of the specimen, since it diminishes the full inner reflection at the interface crystal—air.

On the other hand, immersion also causes the brightness of luminescence to decrease for the following two reasons. In immersion, α-particles travelling between the crystals of the specimen are absorbed before they can produce luminescence. At the same time immersion weakens the «directing action» of a rough surface (Fig. 1). Owing to roughness, radiations from points *A*, *B*, *C*, and so forth, are concentrated in the required direction to a greater extent than would be the

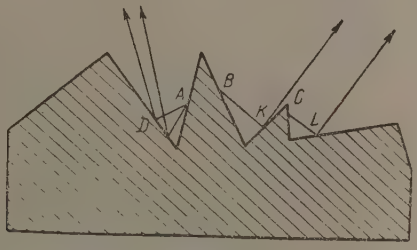

Fig. 1

case if the surface of the powder were perfectly smooth, since the radiation at every point emanates in a greater solid angle.

As has already been pointed out, the brightness of autoluminescence in our specimen increased upon immersion. Hence the first two factors prevail.

As a matter of fact, special experiments with smooth surfaces showed that the increase in brightness upon immersion does not depend upon the roughness of the surface, *i. e.* the «directional effect» does not play an important part. At our request A. G. Samartzeva of the Radium Institute activated PUS with radium. The radium was adsorbed on the surface of the crystals of the salt. The radioluminescence of such a powder was ten times brighter than that of an unactivated sample.

As distinct from the behaviour of the natural autoluminescing specimen, the brightness of the activated sample of PUS decreased upon immersion, as in the case of common zinc sulphide activated with radium. Moreover, for the activated sample, the effect depends upon the state of the surface. When the latter is smooth, immersion in ethyl alcohol results in a decrease in brightness of only 20 per cent, whereas for an ordinary rough surface this decrease is as much as 35—40 per cent. The reason for such a difference in behaviour between the natural and activated samples of PUS lies in the difference in the preferential arrangement of the radioactive atoms in the natural and activated crystals. In the first case the radioactive atoms (of uranium) are located within the crystal, and the α-particles travel mainly inside the active part of the specimen causing luminescence. In the case of the activated sample, on the contrary, the radioactive atoms (of radium) are located on the surface of the crystals, and a considerable fraction of the α-particles travel in the spaces between the crystals. When these spaces are filled up by the liquid upon immersion, a large per cent of the α-particles is absorbed before they can produce luminescence.

The same holds when activated PUS with smooth and rough surfaces is immersed. When the surface is rough, a certain part of the α-particles may travel from the crystal surface to adjacent crystals, *e. g.* (Fig. 1) from *A* to *D*, from *B* to *K*, from *C* to *L*, *etc.*, and produce luminescence there. Upon immersion, how-

ever, these α-particles are absorbed on the way without luminescing. For the natural specimen, the fraction of such α-particles is negligible.

Different liquids affect the activated PUS differently. In the case of a common rough surface, the brightness of luminescence of an activated sample decreases by 35 per cent upon immersion in ethyl alcohol, and by 55 per cent upon immersion in benzene, even though absorption of α-particles in both media is nearly the same. This may be explained by a greater weakening in the «directing action» of the surface in benzene as compared with ethyl alcohol, since the index of refraction of benzene is nearly the same as that of PUS, and the rough surface of the powder covered by benzene becomes smoother from the optical standpoint than when it is covered by ethyl alcohol.

4. Except for minor modifications, the method we employed to obtain potassium uranyl sulphate ($K_2UO_2(SO_4)_2 \cdot 2H_2O$) was essentially the same as that described by Gmelin and Kraut [3]. The crystal size is greatly affected by fluctuations in the temperature of crystallization and the purity of the salt. Thus, when the temperature fluctuated within one degree, we obtained a plate comprised of intergrown crystals 2.5—3.5 mm in length, while when the temperature fluctuation did not exceed 0.2°C, crystals 4—6 mm in length were obtained. The plates comprised of the largest crystals were used to prepare standards of two dimensions: with a circular luminescing surface 10 mm in diameter, and with a rectangular or circular luminescing surface 35 × × 30 and 30 mm in diameter, respectively. To protect the plate of PUS crystals, it was inclosed in a glass case glued together with a special plastic—balsamine, and fitted into a metal protective casing. The brightness of the standards was determined photometrically in collaboration with P. P. Feofilov. For photometering served a Lummer cube observed through a magnifying glass with a six-fold magnification, since the low brightness of the standards requires large angular dimensions of the fields of comparison (~15°) in photometry. One field of the cube was illuminated by the light of a standard lamp of known light intensity and colour temperature, reflected from a barium sulphate screen. The working conditions of the lamp were rigidly controlled. The brightness of the screen was varied by means of a neutral gelatine wedge. The albedo of the screen was taken equal to 0.95. The second field of the cube was illuminated by the measured standard placed close to it. Ten series of measurements consisting of 20 readings each were carried out for every sample by two observers. This was done to obtain greater accuracy, which was about 5 per cent. The mean value for brightness obtained in this manner is $1.2 \cdot 10^{-4}$ apostilbs for the small standards and $2.5 \cdot 10^{-4}$ apostilbs for the large standards.

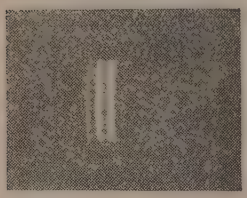

Fig. 2. Photoluminescence and autoluminescence spectra of potassium uranyl sulphate

The brightness of the standards may be increased several fold if the crystals are glued with a colourless spirit varnish on to the inner surface of a hollow sphere that has been preliminarily painted with some white scattering layer, *e. g.*, calcium oxide or magnesium oxide. In this sphere a small opening is made. Brightness of such an opening 10 mm in diameter is some three times that of small standards of the same diameter. Knowing the brightness and the spectrum of the autoluminescence of the PUS, one may compute the absolute energy.

To determine the spectrum, we used a State Optical Institute spectrograph[4] having a relative aperture of 1 : 0.5. Using a wide slit we succeeded in obtaining after a 30-hour exposure the three most intensive bands in the autoluminescence spectrum, which coincided in the position of bands and energy distribution with the fluorescence spectrum of the PUS. This can be clearly seen in Fig. 2, where an enlarged photograph of autoluminescence spectrum (in the centre) and the fluorescence spectrum (on the edges) has been reproduced (the photograph is magnified). The fluorescence spectrum of PUS was studied in detail by Nichols

and Howes ([5]) and Lewschin ([6]). It consists of five bands with maxima at 589, 561, 536, 513 and 492 mμ. The bands at 513 and 536 mμ are the most intense.

Working with the measured value of the brightness and the energy distribution in the fluorescence spectrum, P. P. Feofilov carried out a computation of the energy radiated by 1 cm² of the surface of the standards and obtained $1.1 \cdot 10^{-5}$ erg/sec. A computation made of the energy radiated by 1 cm² of the surface was also made on the basis of a direct photometric comparison of the energy of autoluminescence of PUS with the energy of a lamp standardized against a black body and provided with a filtre that cut out the region of luminescence of PUS. In this way a valence of $1.8 \cdot 10^{-5}$ erg/sec. was obtained. It may be shown that when the energy yield is 100 per cent, the energy radiated by a plate 1 cm² in area should be equal to $1.1 \cdot 10^{-1}$ erg., *i. e.*, the actual energy yield is only 0.01 per cent. The quantum yield is very great, namely, ~335 quanta per α-particle.

5. All the measurements were carried out at 20—25°C, but the standards may be used without any correction at any temperature between −50 and +50°C, since within this interval luminescence is independent of the temperature.

Since the brightness of the standards is very low, of the order of 100 thresholds of the sensitivity of the human eye *, measurements with them are very fatiguing and require preliminary training.

Besides the fact that the standards and fields of comparison must be of large angular dimensions and that a magnifying glass has to be used, it is necessary to completely adapt the eye to darkness (to spend at least an hour in totally dark room). Hence such standards should be regarded as primary standards against which from time to time are checked secondary working standards prepared from radioactive luminescent substances having greater brightness, though gradually changing with time.

In conclusion I consider it my pleasant duty to express my deep gratitude to S. I. Wawilow, at whose suggestion and under whose supervision this work was carried out, and also to V. G. Chlopin and A. G. Samartzeva for their consultation and help in activating potassium uranyl sulphate. I am also extremely grateful to P. P. Feofilov for his friendly help and advice.

The State Optical Institute.

Received
7. IX. 1944.

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* The lowest brightness that the human eye can perceive is of the order of 10^{-6} apostilbs.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PHYSICAL CHEMISTRY

CONCENTRATION ISOTHERMS OF CATALYST POISONING

By S. Z. ROGINSKY, Corresponding Member of the Academy

In a recent paper (1) we described equations connecting the activity of a catalyst with the amount of poison adsorbed. Since experimental measurement of this latter quantity in the course of the process involves great difficulties, it is customary to determine the dependence of the activity upon the equilibrium concentration of the poison. This, however, has no sense unless the poisoning is reversible.

Let us consider a non-uniform surface and denote the distribution of the ctivation energy of the catalytic process over this surface by ρ (E). In deriving the Φ [Γ] isotherm (1), the law of the poisoning process was determined entirely by the type and position of the leading band and by the order in which the different sites were gradually blocked. But in the case of concentration isotherms this simple method is inadequate. The relation between the partial pressure of the poison in the volume π and the magnitude and configuration of the adsorption must be defined more accurately.

A. The simplest in character is poisoning that begins from the least active sites and ends with the most active ones forming the stationary band at the edge of the distribution. Here the rate of the process is affected only by the blocking of the active sites forming the band. This blocking should be purely Langmuirian in character:

$$w=K\Psi\Pi p_i=k_1\Psi\rho_1\Pi p_i \qquad (1)$$

In this equation

$$\Psi=1-\Phi=1-\sum\frac{a_ip_i}{1+\sum a_ip_i}=\frac{1}{1+\sum a_ip_i} \qquad (2)$$

where p_i are the partial pressures of the adsorbed substances, including π of the poison, and $a_i=b_i^{-1}$ is the adsorption coefficient. Resolving $\sum a_ip_i$ into

$$\sum{}^{1}a_ip_i+a_\Gamma\pi=\sum a_ip_i \qquad (3)$$

we reduce equation (2) to the following form:

$$A=\frac{A_0}{C+a_\Gamma\pi} \qquad (4)$$

Thus, under definite conditions of poisoning catalysts with non-uniform surfaces may behave like those with uniform ones.

We note further that in this case the critical boundary which exists with respect to Γ but not to π vanishes completely.

B. In case poisoning begins with the most active points, the picture is essentially different. Here the main effect of the poison is to shift the con-

trol band and to increase the mean activation energy. To derive the concentration isotherms, it is necessary to establish a relation between E_Γ in the equation

$$A_\Gamma = M_\Gamma \exp[-\gamma E_\Gamma / RT] \qquad (5)$$

(where $\gamma = 1$ for power and uniform distributions and $\gamma = 1 - \alpha RT$ for exponential distributions) and the concentration of the poison π, or inasmuch as

$$\pi = b_0 e^{-Q_\Gamma / RT} \qquad (6)$$

i. e.

$$Q_\Gamma = -RT \ln(\pi / b_0) \qquad (6a)$$

between E_Γ and Q_Γ. Equations (6) and (6a) come from our works on the theory of adsorption on non-uniform surfaces ([2-3]). To establish a relation between E_Γ in (5) and Q_Γ use is often made of the linear dependence

$$E_\Gamma = E_0 - \beta Q \qquad (7)$$

which, although it cannot be strictly substantiated, frequently leads to the correct conclusion and is convenient to operate with mathematically. Making use of equation (6a), let us replace E_Γ in equation (7) by:

$$E_\Gamma = E_0 - \beta Q_\Gamma = E_0 + \beta RT \ln(\pi / b_0) \qquad (8)$$

This leads to a general equation:

$$A(\pi) = M_\Gamma \exp(-\gamma E_0 / RT) \exp\left(-\frac{\gamma\beta RT \ln(\pi / b_0)}{RT}\right) = \frac{A_0}{\pi^{\gamma\beta}} = \frac{A_0}{\pi^m} \qquad (9)$$

which holds for all types of ρ, when (7) may be used.

The result is highly instructive. It shows indeed that in the most typical case, when blocking begins with the most active points, the character of the dependence of the activity upon π is exactly the same for all distributions. When the relation between E and Q is linear, it leads to the universal equation:

$$\lambda \simeq \pi^{-\gamma\beta} = \pi^{-m} \qquad (9a)$$

which, unfortunately, tells us nothing about the distribution type. Equations (9) and (9a) differ from the Langmuir poisoning equations by the exponent of π $\gamma\beta = m$, which in the general case is not equal to -1 and may assume any fractional values.

C. More unique is the character of the concentration isotherms in the case of a mobile «suspended» band ([1]). During the action of a contact poison on such a distribution, the phenomenon of critical poisoning quantities changes into the phenomenon of critical concentrations. Up to a certain Φ_Γ there is practically no poisoning; when the poisoning front reaches the control band, the latter is set in motion and the activity begins to fall abruptly. The equation describing the fall will depend upon the direction of the motion. If the poison moves from the direction of small values of E, equation (9) becomes applicable further on. If it moves from the large E, the equation is somewhat different in view of the fact that E and Q_Γ move in the same sense, and that the heat of formation of the transition complex Q^* enters it. Let us assume for simplicity that

$$E_\Gamma = E_0 + \beta_1 Q^* \qquad (7a)$$

and at the same time that

$$E_\Gamma = E_0 + \beta_2 Q_\Gamma \qquad (7b)$$

As the band shifts, the activity changes in accordance with the equation

$$A(\Gamma) = M_\Gamma \exp(-\gamma E_\Gamma / RT) \exp(+Q^* / RT) = M_\Gamma \exp\left[-\frac{\gamma E_\Gamma - \gamma' E_\Gamma + \gamma' E_0}{RT}\right]$$

or

$$A_\Gamma = M_\Gamma \exp(-\gamma' E_0 RT) \exp(-\gamma^* E_\Gamma / RT) \tag{10}$$

where $\gamma' = 1/\beta_1$ and $\gamma^* = \gamma - \gamma'$.

Substituting

$$E_\Gamma = E_0 + \beta_2 Q_\Gamma = E_0 - \beta_2 RT \ln(\pi/b_0) \tag{8a}$$

we get:

$$A(\pi) = A_0 \pi^{\gamma^* \beta_2} = A_0/\pi^m \tag{11}$$

where $m > 0$.

The condition for the appearance of a suspended band is that $\gamma^* < 0$, since otherwise sites with small E and small Q would form a control band.

Hence, in this case, too

$$A = A_0/\pi^m, \quad \lambda \simeq \pi^{-m} \tag{9b}$$

where

$$m = -\gamma^* \beta_2 = (\gamma - \beta_1^{-1})\beta_2 = (1 - \alpha RT - \beta_1^{-1})\beta_2 = (\beta_3 - 1)\beta_2 \tag{9c}$$

$\Phi[\Gamma]$ isotherms of poisoning are often straight lines. This is more seldom the case with reversible poisoning, since the most typical instances of linear isotherms are kinetic in nature. That they sometimes do occur in the latter case, too, is due to the lack of correlation between the activity of the site and the heat of adsorption of the poison on it.

In this case the relation between Γ and π is determined by the type of distribution with respect to Q_Γ, and the concentration isotherms are obtained by the simple substitution for Γ of its isotherm ($a\pi^{1/n}$ in the case of exponential distribution, $a \ln \pi$ in the case of uniform distribution, *etc.* [2]).

It is not clear whether poisoning isotherms of the type

$$\begin{aligned} \lambda &\simeq 1 - a^1 \pi^{1/n} \\ \lambda &\simeq 1 - a^1 \ln \pi \end{aligned} \tag{12}$$

etc., are actually met with.

The results presented above may also be obtained more directly, but somewhat less clearly, by proceeding from the theory of the adsorption of mixtures developed recently by the author of this paper and O. M. Todes for distinctly non-uniform surfaces [2].

Received
20. III. 1945.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

GEOCHEMISTRY

HYDROCHEMICAL FACIES OF LAKE (AND SEA) WATERS

By G. A. MAXIMOVICH

(Communicated by A. A. Grigoriev, Member of the Academy, 28. VIII. 1944)

Lakes are an important variety of the hydrosphere. Their area makes up about 1.8 per cent of the dry land, or 2 682 000 km², and the volume of water in them is 250 000 km³. This constitutes 0.75 per cent of the area of the ocean and only about 0.02 per cent of its volume.

The hydrochemistry of lakes is of great interest. The amount of mineral substances dissolved in the lakes varies from $6 \cdot 10^{-5}$ to 37.15 per cent, that is the concentration of the lake waters varies $6.19 \cdot 10^6$ times ([2],[6],[7]).

The water filling the lake bath is not uniform in composition. Aside from silt waters, which should be classed among underground waters, the lakes always contain no less than two varieties—bottom and surface, or basic waters. Their gas and mineral composition is different.

The fact that the composition of the prevailing substances dissolved in lake (or sea) water varies according to area and depth, made the author suggest the conception of the hydrochemical facies of lakes (and seas).

A hydrochemical facies of a lake (and sea) is such a portion of the basin, whose waters are characterized by definite hydrochemical conditions due to the prevalence of definite substances dissolved in them (ions, colloids). The concentration and mineral composition of waters within each facies may vary, but the predominance of a given substance characterizing it is maintained.

Like in river waters ([4]), the hydrochemical facies of lake (and sea) waters is determined from three components predominant by weight: they receive their names in the order of their decreasing value.

The hydrochemical facies are united into groups or formations, according to the first predominant dissolved component. From 588 analyses of waters of 209 lakes (and seas) taken from the literature ([2],[3],[5–7]), the author separated out 43 hydrochemical facies, according to the concentration of their waters. They are given in the table.

The lake (and sea) waters may be grouped into five hydrochemical formations: the siliceous, hydrocarbonate, sulphate, sodium and chloride one.

The distribution of the hydrochemical facies of lake waters over the Earth surface is characterized by a latitudinal and altitudinal zoning. The following zones may be distinguished:

1. Sone of the predominance of silica and hydrocarbonate-silica hydrochemical facies of tropical and subtropical lakes.

2—3. Zones of the predominance of sulphate, sodium, chloride, hydrocarbonate-sodium and hydrocarbonate-potassium hydrochemical facies of lakes of the northern and the southern hemisphere, confined to the desert-steppe zones.

4—5. Zones of the predominance of hydrocarbonate-calcium hydrochemical facies of lake waters of the temperate climatic zone of the northern and the southern hemisphere.

Hydrochemical Facies of Lakes (and Seas)

Groups of facies (formations)	Facies *	Mineralization in 0.0001%	Number of lakes	Number of analyses	Lakes (and seas)
Siliceous	Si-H-Na	80—118	2	2	Yellowstone and Crater Lake (Oregon), USA
Hydrocarbonate	H-Si-Ca	73	1	1	Lake Tahoe (Cal.)
	H-Ca-Si	16—118	6	34	Lake Superior, Lake Huron, Lake Michigan, Minetonka Lake, Rangeley (Main), Chempeks, Alps
	H-Ca-S	14—272	22	51	Baikal, Kuzher, V. Muzhinskoe, S. Muzhinskoe, Babit (Latvia), Erie, Mushid (USA), Winnipeg, Red Lake (Canada), Zürich, Lac Léman, Taney, Amaldigen (Alps), Königssee, Walchensee, Schliersee, Chiemsee, Küchelsee, Tegernsee (Germany)
	H-Ca-M	106—178	6	7	Lake of Pskov, Angernsee (Latvia), Würmsee (Germany), Lake of Annecy (France), Lago di Garda (Italy), Mille Lacs (USA)
	H-Ca-Ch	137—155	5	5	Surok, Kozhelaier, East Yalchevskoe, Halstedsee (Germany), Okeecho-bee (Florida)
	H-Ca-Na	39—294	9	9	Okunevoe, Surok, Kuzher, Urdur, M. Muzhinskoe, Kozhlaier, Puzhanier, Yalchevskoe (main), Baikal (USSR)
	H-Na-M	315	1	1	Kisykul
	H-Na-Si	379	1	1	Silver Lake (Oregon)
	H-S-Ca	160—512	3	10	Plattensee (Hungary), Winnipeg, Red Lake (Canada)
	H-Ch-S	304.344	1	1	Svinoe (Budenovsk)
	H-Ch-Na	49—3374	7	7	Onega Lake, Mayen, Uruskul, Kainkul, Karagaikul, North Kalinovskoe, B. Chebachié (USSR)
	H-Na-Ca	2966	2	2	Gusinoe, Moses Lake (Washington)
	H-Na-K	51.700	1	1	Filan (Nebraska)
	H-Na-S	484—5704	7	7	M. Tzagan-nor, B. Kuzherskoe, Gudzhirnoye (lakes of Oregon and Washington)
	H-K-Na	27300—71200	3	3	Lakes of Nebraska
	H-K-S	53.600	1	1	Fleid (Nebraska)
	H-Na-Cl	135—103470	5	5	Victoria-Nyanza, Natron Lake (Egypt), Kaldy, Kezhkul (N. Kazakhstan), Good-noon (British Columbia)
	Total 17 facies	14—103470	81	146	

Groups of facies (formations)	Facies*	Mineralization in 0.0001%	Number of lakes	Number of analyses	Lakes (and seas)
Sulphate	S-H-Si	654	1	1	Big Stone Lake (Minnesota)
	S-H-Ca	224—306	2	2	Utah Lake (Utah), Winnipeg (Canada)
	S-Ca-H	122—2335	5	9	Nouar, Rietom (Alps), Vad (near Penza), Goluboye (Tataria), Kungur (in a cave)
	S-Ca-M	2373	1	1	Rietom (Alps)
	S-Ca-Na	4446	1	1	Chichankanab (Yukatan)
	S-Ch-Na	1165—40626	4	8	Tagar (Siberia), Pilenkino (Azov), Algren-sor, Utah (Utah)
	S-Na-H	48360—353700	2	4	Gudzhirnoe, Beloe (Siberia)
	S-Na-Ch	11278—145500	15	16	Pilenkino (Azov), Gudhirganskoe, Turpanye, Obrochnaya lestchad, lakes of Abagan, Wyoming, North Dakota, Canada, Rumania
	S-Na-M	6708	1	1	Wyoming
	S-M-Ch	193500	1	1	Muskoka (Canada)
Total 10 facies		122—353700	33	44	
Sodium	Na-H-S	Per cent	1	1	Winnipeg (Canada), azonal
		1.85—11.97	5	5	Lakes of Peru, Nebrasca, Washington, Wyoming, California
	Na-H-Cl	0.09—5.12	5	7	Boga-Khan-Nor, Verkhn. Beloe (Siberia), lakes of Oregon and California
	Na-S-H	0.63	1	1	Riszanda (Hungary)
	Na-S-Ch	0.49— 0.61	2	2	Selenga, Tulare Lake (California)
	Na-Ch-S	0.25— 1.51	2	3	Kirenskoe (Siberia), Walker Lake (Nevada)
	Na-Ch-H	1.05—21.37	5	9	Van, Harney Lake, Albert Lake (Oregon), California lakes
Total 6 facies		0.09—21.37	21	28	
Chloride	Ch-Ca-Na	0.366	1	1	Charkhal (N. Kazakhstan)
	Ch-H-Na	5.336	1	1	Chebarkul (USSR)
	Ch-Na-H	0.09—21.14	8	10	Maloe Chebachie, May-Balyk (N. Kazakhstan), Lakes of Nevada, India, South Africa
	Ch-S-Na	0.089—28.9	6	8	Aral, Tinetzkoe, B. Manych, Kirghizia, Rybnoe, Charkhal (N. Kazakhstan)
	Ch-Na-S	1.11—31.0	39	316	Atlantic Ocean, Indian Ocean, Arctic Ocean, White Sea, Baltic Sea, Ireland, Black, Mediterranean, Red, China Seas. Lakes: Kuku-Nor, of Iran, the Caspian, Kara-Bugaz, Iletzkoe, Tagarskoe, Alatyr, B. Leshchad, Selenga, Lomovoe, B. Balpash, Balpash, Basaybar, Moldybay-sor, Chuiruk-sor, Taykonur, Bish-tuz, B. Selenoe (Utah), Chile, Argentine, Africa, India
	Ch-S-M	—	1	1	Kara-Bugaz
	Ch-Na-M	4.6—37.15	13	19	Khanskoe, B. and M. Jarovoe, B. and M. Kalkamak, Kammertauz, Ekibastuz, Altybai-sor, Dead Sea, Sivash, Inder, Lakes of Rumania, Australia
	Ch-M-Na	19.26—26.5	2	9	Dead Sea, Elton
	Ch-M-Ca	25.99	1	1	Dead Sed, 300 m (Palestine)
Total 9 facies		0.09—37.15	72	366	
Total 43 facies		14.10^{-4}—37.15	209	586	

* Si — siliceous, H — hydrocarbonate, Na — sodium, Ca — calcium, S — sulphate, M — magnesium, Ch — chloride, K — potassium facies.

6. Mountain (vertical) zone of the predominance of silica and hydrocarbonate-silica hydrochemical facies of lake waters. The waters of black and brown lakes of near-polar and tropical regions ([2]), rich in organic matter, belong to the silica-carbonate (SiO_3—C) and hydrocarbonate-carbonate-silica (HCO_3—C—SiO_2) facies.

The hydrochemical facies of lake waters is a typical complex of predominating solutes pointing to definite climatic and, accordingly, geochemical (weathering), soil, hydrogeological and hydrobiological conditions under which lake waters acquired their concentration and composition.

The local lithological, hydrological and hydrogeological conditions, as well as the activity of man, may cause the appearance of azonal hydrochemical facies of lakes. Zonal facies are predominant, however.

The development of readily soluble rocks causes the appearance of azonal hydrochemical facies. The karst lakes of the Kungur region, inclusive of the underground lakes in the Glacial cave ([3]), within the zone of the predominance of hydrocarbonate-calcium facies, are characterized by azonal sulphate-calcium-hydrocarbonate facies. Thus, to this facies in the Kazan region should be referred lake Goluboye (Blue), and in the region of the town of Penza, lake Vad.

Lake Winnipeg, located within the zone of the predominance of hydrocarbonate-calcium waters of the northern hemisphere, is fed by the Red River. This river is fed in the zone of sulphate and chloride facies, and the lake near the point of discharge of the river belongs ([7]): in July, to the azonal SO_4—HCO_3—Ca facies; in June, to the Na—HCO_3—SO_4 facies; in August and October, to the transitionary HCO_3—SO_4—Ca facies, and in September only, to the zonal HCO_3—Ca—SO_4 facies. Further off the mouth of the river, the lake is characterized during the year by the zonal HCO_3—Ca—SO_4, or the transitionary HCO_3—SO_4—Ca facies. Azonal phenomena are also observed ([1]) in such cases where a considerable rôle in the supply of the lakes is played by mineral springs.

The hydrochemical facies change with time and space. With a change in the climatic conditions the boundaries of the zones of hydrochemical facies are shifted. The hydrochemical facies of lake waters at the point of discharge of rivers into the lakes may vary in the course of the year, according to the condition of the river waters supplied. The activity of man may also bring about changes in the hydrochemical facies of a lake by controlling water supply from rivers. Such phenomena have been recorded for Lake Utah ([6]).

The hydrochemical facies of lake waters change both according to the area of lakes (and seas), and along the vertical. Thus, the lakes may be monofacial and polyfacial. The former type is more common to lakes than to rivers, and chiefly to small lakes. The polyfacial type is common to more concentrated lakes, into which discharge rivers fed from another hydrochemical zone.

The waters of seas and oceans belong to the chloride-sodium-sulphate facies. This monofacial character of the World ocean (the slightly mineralized Baltic Sea and the highly concentrated Red Sea included), alternating with transitional facies at the points of river discharge, is due to the fact that the concentration of the waters of the Ocean varies within a rather narrow range (0.7—6.15 per cent). This concentration is characterized by the chloride-sodium-sulphate facies.

Received
28.VIII.1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

BIOCHEMISTRY

BACTERIOSTATIC SUBSTANCES OF ANIMAL ORIGIN

By A. S. KONIKOVA, A. P. URASOVA and R. M. ASARKH

(Communicated by J. O. Parnas, Member of the Academy, 5. IX. 1944)

According to Dubos' idea [1] of the evolution of bacteria as antagonists of pathogenic microbes, it seemed likely that animal evolution might also have given rise to such antagonistic agents. We therefore employed a method similar to that developed by Dubos for the preparation of tyrothricin, with the aim of obtaining active bacteriostatic substances from animal tissues. We have succeeded in obtaining from rabbit liver and human placenta substances producing bacteriostatic effect upon Gram-positive microbes, such as staphylococci, streptococci, *Bac. brevis*, *etc.*

M e t h o d. Fresh rabbit liver or human placenta were minced in Latapi mill, and the mince was extracted during 24 hours at 13—15° C with two volumes of *m*/15 phosphate buffer (pH = 7.8). The extract was filtered through double gauze, and concentrated hydrochloric acid was gradually added until complete precipitation of the proteins ensued. The bulky precipitate was collected of filters and extracted during one day with a threefold amount of acetone. The acetone extract was filtered and concentrated *in vacuo* until the solution became opalescent. The condensed extract was poured into 10 volumes of 1 per cent NaCl solution, or 5 volumes of 1 per cent HCl, dried, washed with ether and kept in a vacuum-desiccator with P_2O_5. The dried preparation is a dark friable, or somewhat sticky mass.

The bacteriostatic activity was determined as follows [2]: 20 mg of the preparation were dissolved in 1 ml of 95 per cent ethyl alcohol. Varying amounts of this solution were mixed in a test-tube with 10 ml nutrient agar, and the mixtures poured into Petri dishes. The concentration of alcohol in the agar did not exceed 1 per cent. Inoculates of an 18-hour culture of bacteria were grown at 37°C on solid agar containing different amounts of the preparation. Corresponding volumes of the solvent (alcohol) were added to control plates. Readings were taken after 24 hours' incubation (Table 1).

No bacteriostatically active fractions were obtained from bovine or rat liver, rabbit bile, blood or muscles.

60—80 mg active substance were secured from 1 kg placenta and 500—600 mg from 1 kg liver. The bacteriostatic preparation from liver is soluble in alcohol and acetone, and water-insoluble. Different preparations were found to contain 1.3—1.9 per cent nitrogen (Kjeldahl method). Liver preparations kept at room temperature lose much of their activity after 2—3 weeks. In alcohol solution the rate of inactivation is higher.

Boiling during 10 minutes does not affect the activity of the preparation. Washing with ether does not alter essentially the activity per unit dry weight.

Table 1

Bacteriostatic Activity of Liver and Placenta Preparations
(bacterial growth in γ per 1 ml agar)

Species of bacteria	200	100	50	25	10	0 (control)
With liver preparation						
Staphylococcus aureus	—	—	—	++	++++	++++
Streptococcus haemolyticus . .	—	—	—	—	+	++++
Bacillus brevis	—	—	—	++	+++	++++
Echericha coli	++++	++++	++++	++++	++++	++++
With placenta preparation						
Staphylococcus aureus	—	—	—	+	+++	++++
Streptococcus haemolyticus . .	—	—	—	+	+++	++++

Note. Growth, +; no growth, —.

Trypsin does not destroy the active principle of the liver preparation; on the contrary, it restores the activity deteriorated through prolonged storage. The inactivation of the liver preparation may be possibly caused by some decay products of contaminating proteins. Upon a prolonged trypsin treatment (3—4 days) of the bacteriostatic liver fraction in 0.4 per cent $NaHCO_3$, a precipitate is formed which retains its bacteriostatic activity.

The effect produced by the liver preparation upon some enzymic processes in susceptible microbes is similar to that of tyrothricin isolated from soil bacteria.

We have investigated the effect of the liver preparation upon glucose dehydrogenase in *Staphylococci* and *Bac. brevis* (Table 2).

Table 2

Effect of Liver and Placenta Preparations upon Activity of Glucose-Dehydrogenase in *Staphylococci*
Composition of the sample: 1 ml washed bacterial suspension + 1 ml 10 per cent glucose + 1 ml *m*/15 phosphate buffer (pH 7.2) + 1 ml 0.02 per cent methylene blue + varying amounts of the preparation

Amount in γ/ml	Liver				Placenta				
	200	100	50	25	200	100	50	25	Control
	N	N	N	PR	N	N	PR	PR	CR

Note. N — no reduction; PR, partial reduction; CR, complete reduction.

From the above data it may be seen that the substance isolated from liver is similar to tyrothricin in completely inhibiting dehydrogenation of glucose by *Staphylococci*, somewhat higher concentrations of the former being required, in accordance with its lower bacteriostatic effect.

The amino-acid metabolism of *Bac. brevis* is affected by the bacteriostatic fraction of the liver. It inhibits (completely in concentrations as low as 100—50 γ/ml) the dehydrogenation and deamination of asparagine, *l*-glutamic acid, *d*- and *l*-alanine and *l*-valine. Tyrothricin was found to produce a similar action. Deamination was evaluated by the formation of ammonia (Conway's method); dehydrogenation, by the rate of reduction of methylene blue (Thunberg) (Table 3).

Table 3

Effect of Bacteriostatic Preparations on Deamination and Dehydrogenation of Amino-acids and Asparagine.
Deamination. Composition of sample: 10 μ mol of amino-acid + 1 ml washed bacterial suspension + 2 ml *m*/15 phosphate buffer (pH 7·2) + 3.0 — 0.75 mg liver preparation or 3.0—0.35 mg tyrothricin. 2 hours incubation at 37° C.
Dehydrogenation. Composition of sample cf. Table 2

Substrate	Enzymic process	Substrate + +bacteria	Substrate + bacteria + liver preparation	Substrate + bacteria + tyrothricin preparation
Glutamic acid	Deamination	58	O	O
»	Dehydrogenation	CR	N	N
Asparagine	Deamination	85	O	O
»	Dehydrogenation	CR	N	N
l-valine	Deamination	41	O	O
»	Dehydrogenation	CR	N	N
l-alanine	Deamination	52	O	O
»	Dehydrogenation	CR	N	N
d-alanine	Deamination	48	O	O
»	Dehydrogenation	CR	N	N

Note. The extent of deamination is expressed in per cent to the amount of the substrate added. N, no reduction; CR, complete reduction.

It appears from our data that bacteriostatic substances do occur in some organs endowed with a barrier function, such as the liver or the placenta.

These substances can be isolated by a procedure similar to that used for extracting tyrothricin from soil bacteria. This bacteriostatic effect is accompanied by the inhibition of glucose-dehydrogenase and of amino-acide dehydrogenase of the susceptible bacteria.

The purification and identification of the active principle is in progress.

Received
5. IX. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

CHEMICAL TECHNOLOGY

OBTAINING OF PORCELAIN WITHOUT FELDSPAR FROM TURKMENIAN RAW MATERIALS

By M. A. BESBORODOV

(Communicated by D. S. Beliankin, Member of the Academy, 12. VIII. 1944)

The classical mixture for porcelain body is known to be 50 per cent clay, 25 p. c. quartz and 25 p. c. feldspar ([1—3]). Of course, this ratio of the ingredients may vary within certain limits depending upon the intended use of the manufactured article, the temperature of firing and the technological process. But in any case all three are always present in the usual porcelain body.

The clay ingredient the most commonly used is kaolin. To make the body suitable for forming, a plastic fire clay containing minute quantities of oxides of iron and titanium is added besides. Feldspar acts as a flux, and it too should contain slight quantities of colouring oxides.

Before the war Turkmenia and Central Asia as a whole did not have a porcelain industry. For many years all porcelain ware was imported from factories in the European part of Russia. In 1942 research on the production of porcelain from domestic raw materials was started under the supervision of the author at the Turkmenian Branch of the Academy of Sciences of the USSR. Since kaolin and feldspar do not occur in Turkmenia, a search had to be made for other materials capable of yielding a normal porcelain product. After trying out systematically the available raw materials, we settled on a mixture not containing feldspar that enabled us to obtain on March 1, 1943, in Ashkhabad, the first specimens of porcelain of the requisite whiteness and a characteristic lustrous conchoided fracture.

Its main ingredient is a clay of finely schistous structure from the Djardanak deposit, which gives a white product after firing. Bentonite from the Djebel, deposit acts as the flux, partly fulfilling the rôle of plastifier as well. The third component is the plastic fire clay from the Vandob deposit, which serves mainly as plastifier. All these materials are now being obtained for industrial purposes. A chemical-technological investigation of the Djardanak clay, of which there has been no mention until now in the literature, was undertaken by us in 1942. We found this clay to possess highly interesting qualities both from the scientific and industrial standpoints.

The Djardanak clay deposit is situated in the Kugitang-Tau Mountains, in south-eastern Turkmenia. It lies not far from the village of Djardanak, 45 km to the North of Boldyr Station on the Ashkhabad railway line. According to preliminary survey data of the Turkmenian Geological Office, its reserves amount to 10 000 tons. However, there is reason to believe that they are actually much greater. In the same neighbourhood, some 7—12 km away from the Djardanak deposit, lies the Vandob clay deposit, near the village of Vandob from which they derive their name. This deposit has been estimated by the Turkmenian Geological Office to run into 50 000 tons. Near Djebel, a station on the Ashkhabad railway line, there is a bentonite deposit known by the name of Oglanly. The resources of bentonite here are estimated at 1 500 000 tons.

In Table 1: 1) the analysis data in per cent for a specimen of Djardanak clay, 2) are compared with those for Zetlitz kaolin, which is the world standard in porcelain manufacture, adopted at the international conference on inorganic chemistry in 1924, and with 3) the technical standards for fine ceramic industry kaolin as specified by the All-Union Standards' Committee [OST] No. 2371.

Table 1

Chemical characteristics	Djardanak clay	Zetlitz kaolin	OST 2371 Grade I	II	III	IV
Fe_2O_3	0.45	0.43— 1.01	0.6	0.8	1.0	1.5
TiO_2	0.19	0.36— 0.50	0.6	0.8	1.0	1.2
CaO	0.37	0.08— 0.78	0.8	0.8	0.8	0.8
Al_2O_3	39.27	37.19—39.60	—	—	—	—
CaO + MgO + alkali	3.98	0.75— 2.86	—	—	—	—
SiO_2	44.17	45.66—47.83	—	—	—	—

It is obvious from this table that the Djardanak clay is a very good material for making porcelain. Its high alkali content makes it an excellent ingredient for a porcelain body devoid of feldspar. Complete analyses of Djardanak clay, as well as Vandob clay and Djebel bentonite, are given in Table 2 (in per cent).

Table 2

Oxides	Djardanak clay		Djebel bentonite		Vandob clay	
	in raw substance	calcined	in raw substance	calcined	in raw substance	calcined
SiO_2	44.17	50.40	64.76	78.31	56.68	63.50
TiO_2	0.19	0.22	0.16	0.19	0.85	0.95
Al_2O_3	39.08	44.40	11.87	14.40	26.65	29.70
Fe_2O_3	0.45	0.51	1.07	1.30	1.20	1.32
CaO	0.37	0.42	0.56	0.69	0.50	0.56
MgO	0.20	0.23	2.64	3.22	0.78	0.88
K_2O	2.76	3.15	0.32	0.39	2.26	2.53
Na_2O	0.65	0.74	1.22	1.50	0.50	0.56
H_2O at 110°	1.22	—	13.28	—	2.05	—
Loss on ignition	10.98	—	4.30	—	8.67	—

The chemical analyses of Vandob clay and Djebel bentonite show that neither of them contain any quantity of colouring oxides. However, to obtain a white product, the content of these components in the porcelain body should be as small as possible.

The mineralogical composition of the ingredients of our porcelain, as determined by D. S. Beliankin and V. P. Petrov, is presented in Table 3 (in per cent).

Table 3

Raw materials	Kaolinite	Hydromica monothermite part	Muscovite	Quartz	Monothermite	Hydrobiotite, iron oxides	Organic matter	Montmorillonite
Djardanak clay	10—25	70—80	2—3	3—4	—	—	3—5	none
Vandob clay	20—30	—	—	20—25	40—50	3—5	5—7	»
Djebel bentonite	—	—	—	—	—	—	2—3	main mass

According to chemical-mineralogical analysis, Djardanak clay belongs to the category of non-plastic schistous clays of the type of American flint clays, which, however, differ from the latter by their extremely small iron and titanium con-

tent and high alkali content. Its refractoriness is about 1770° C; the sintering temperature is between 1150 and 1200° C. Its natural colour is grey, and when freshly broken is almost black; after firing it is pure white, without the slightest colour shades whatsoever.

Vandob clay might be called «Turkmenian chassovyar clay» in view of its proximity to the Chassovyar clay both in chemical and mineralogical composition. Its refractoriness is 1650—1680°C. In the natural state its colour varies from light to dark grey; after firing it is grey tinged with yellow. Its plasticity as determined by the method of P. A. Zemyatchensky is 5—6.

Characteristic of Djebel bentonite is its white colour, high plasticity and great dispersion. 96.73 p. c. of the bentonite consists of particles less than 0.01 mm in diameter, and 61.76 p. c. are particles less than 0.001 mm in diameter. The melting point of different specimens varies between 1230 and 1300°C, which enabled us to use it in our porcelain as a flux instead of feldspar.

Thus, the ingredients of our porcelain in contrast to those of the classical mixture, are: 1) a schistous clay of the «flint clay» type yielding a white product on burning, 2) an easily melting bentonite clay, and 3) a plastic fire clay of the «chassovyar» type.

A typical composition for such a porcelain is 80 p. c. Djardanak clay, 10 p. c. bentonite and 10 p. c. Vandob clay. The firing temperature is 1300—1320°C.

A study of a large number of mixtures containing different ratios of the above-mentioned ingredients showed that: 1) Djardanak clay increases whiteness and elevates the firing temperature, 2) bentonite lowers the firing temperature and enhances the translucence, and 3) Vandob clay intensifies the grey tint in the final product. This latter circumstance has now induced us to direct our efforts toward eliminating Vandob clay from our mixture.

The results of a chemical analysis of one of the specimens of our porcelain, made by L. B. Tumilovich in the Geological Institute of the Academy of Sciences and analytical data for other specimens are given in Table 4 (in per cent).

Table 4

Oxides	Porcelain free from feldspar, Turkmenian	Technical porcelain manufactured by the Lomonosov works	Porcelain manufactured by the Dulevo works	Sèvres porcelain
SiO_2	58.38	61.12	70.40	59.61
TiO_2	0.65	36.92	25.24	32.50
Al_2O_3	35.02			
Fe_2O_3	0.97	0.72	0.42	0.54
CaO	1.38	0.96	0.35	4.50
MgO	0.55	0.12	0.10	0.19
K_2O	1.89	not determ.	2.17	1.90
Na_2O	0.86		1.32	0.76

A comparison of these analyses shows that the feldspar free porcelain closely approaches common specimens in chemical composition. It differs from some of them by its large alumina content and smaller silica and alkali contents. This decreases its translucence, but increases its thermal endurance. Its whiteness is 68 p. c. as compared with a standard barite plate. A microscopic study of it made by D. S. Beliankin and V. P. Petrov showed it to consist of mullite glass (93—94 p.c.) and quartz grains (6—7 p. c.). Its porosity (closed) is less than 1 p. c. In hardness it surpasses steel, since the latter does not scratch it.

Turkmenian Branch
of the Academy of Sciences of the USSR.

Received
12. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

GEOLOGY

ON CERTAIN FORMS OF OCCURRENCE OF «ALGAI» DOLOMITES OF THE MIDDLE CAMBRIAN IN THE TRANS-ANGARA REGION

By E. A. NECHAEVA

(Communicated by V. A. Obruchev, Member of the Academy, 23. VIII. 1944)

In the course of geological exploration in the Trans-Angara region (the basin of the Ida and Ossa rivers, the right-hand side tributaries of the Angara) the author discovered original forms of occurrence of the deposits of the upper portion of the Middle Cambrian, which are described in the present paper.

The oldest deposits within this territory are represented by a limestone-dolomite complex of the Middle Cambrian (Angara series) subdivided into six horizons from lithological evidence and findings of fauna and flora.

1. Buret horizon: uneven-platy brown-grey dolomites and dolomitic limestones with remains of a fauna of trilobites (*Bathyuriscus obrutchevi* Lerm., *Ptychoparia* or *Solenopleura* and *Anomocare leave* Ang., according to a preliminary determination of E. V. Lermontova). Subordinate to these are intercalations of foliated calcareous shales and massive black cloddy dolomitic limestones with traces of annelides on the bedding surfaces. In the middle portion of the horizon there is a pisolitic layer. Throughout the horizon flint intercalations are widely developed, passing along the strike into flint-dolomitic breccias. Thickness, 125 to 130 m.

2. Horizon of banded dolomites: banded dolomites thin- and medium-platy, at places gypsiferous with bands of finely cavernous oolitic limestones and flint lenses. Thickness, 30 to 40 m.

3. Bituminous horizon: grey and black stinky marlaceous dolomites with intercalations of foliated shales and white marls. In the lower portion there are a bed of fine-grained greenish-grey sandstone and some gypsum lenses. The upper layers are represented by light-coloured limestones with intercalations of flints and a coarsely shelly texture (algal?). Thickness, 30 to 40 m.

4. Horizon of cavernous dolomites: massive and thick-platy cavernous dolomites and breccia, at places spotted. Thickness, 10 to 30 m.

5. Algal horizon: white and grey dolomites with abundant algae and lenses of red and grey flints. In some intervals large caverns and stylolites. Thickness, 70 to 90 m.

6. Horizon of light-coloured marls: yellow and greenish-grey arenaceous marls and areno-marlaceous dolomites. At the base of the horizon there are thick lenses of gypsum. Thickness, 20 m.

The marls of the 6th horizon are quite conformably overlain, with a gradual passage, by bordeaux-red and brown marls and sandstones of the Verkholensk series of the Upper Cambrian, with an apparent thickness of 350 to 400 m.

The widest area of development has the 5th horizon of the Angara series named the algal one, in the main composed of fine-grained uneven-flaty banded dolomites, white and light-grey, less frequently dark-grey, with lenses and intercalations of flints and white marls. Its thickness is 70 to 90 m. At the base of the

horizon lies a bed of massive white marble-like dolomite, facially passing to variegated likewise massive cavernous dolomites. In some intervals there are abundant voids resulting from leached small crystals of gypsum. Along the strike the dolomites of the algal horizon are repeatedly superseded by variegated syngenetic breccias, occasionally attaining a thickness of 10 to 15 m.

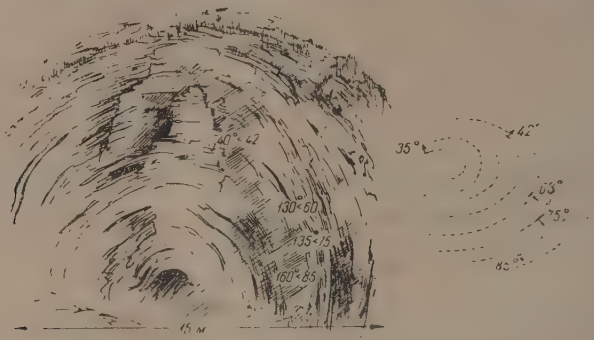

Fig. 1. General view and plan of an individual dome in the valley of the Ida River

The horizon is distinguished by a high content of algae, mainly silicified, of which two types are distinguished.

1. At the base and in the upper portion of the horizon there are found isolated algal forms of a cylindrical, conical and loaf-like shape, showing in the cross-

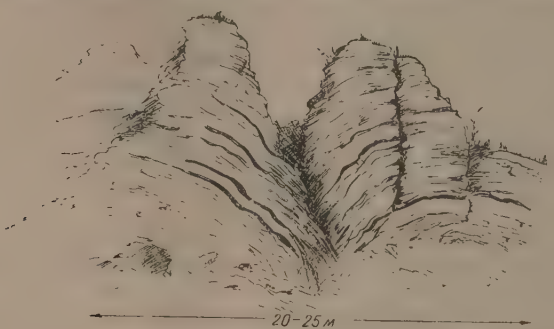

Fig. 2. Juncture of two adjacent domes in the valley of the Ida River

section a concentric-shelly structure. Their dimensions vary from some centimetres to 0.5 m across.

2. In the middle portion of the horizon predominate small columnar algae of the *Collenia* type, growing together with their lateral walls and occasionally composing the whole of the bed.

The texture of the upper part of the algal horizon is highly characteristic. Widespread here are dome-like forms of a shelly structure attaining 30 m across and 10 to 15 m in height (Fig. 1)*. In some areas are observed two or three beds with such texture, the domes of the over- and underlying beds being arranged in a staggered fashion, the core of each dome from the overlying bed lying (along the vertical) on the line of juncture of the two domes of the underlying bed. The angles of incline of the layer, near the cores of the dome-like structures, vary within the range from 30 to 45°. On the periphery they most become very steep, particularly so at the base of the domes, attaining here 70 to 80 , or more rarely

* The drawing was made by N. K. Brimmer.

become gentler, reaching angles of 20—15°; adjacent domes are connected by gentle downwarpings of the synclinal type, or else their steeply-dipping layers make acute angles at the junctures (Fig. 2).

In some structures of this type the dip on the periphery becomes vertical and passes in their lower portion to a reverse dip, thus producing a peculiar spherical shape.

In a number of cases on the limbs of some individual domes minute contorsion is observed, which is not widespread. The layers underlying and covering the beds with a dome-like texture lie practically in a horizontal attitude, or with very small angles of incline.

The dome-shaped structures described on the whole resemble algal reefs, differing from the latter by the fact that the breccias occurring in association with them do not form a belt around the domes, but are located mainly inside them. Such structures have a regional development in the basin of the Ida River over an area of the order of some hundreds of square kilometres.

Received
23. VIII. 1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

GEOLOGY

ON THE PROBLEM OF YOUNG TECTONIC MOVEMENTS IN THE NORTH OF SIBERIA

By V. N. SAKS

(Communicated by V. A. Obruchev, Member of the Academy, 7. VIII. 1944)

The problem of young movements in the northern part of Siberia has long since attracted the attention of investigators. An interesting material in this connexion is furnished by the Taimyr depression—a hollow filled with loose Mesozoic and Cenozoic sediments, wedged inbetween the Hercynian folded structures of the Taimyr peninsula and the Siberian platform composed of Palaeozoic deposits. Along the borders of the depression faults are present, forming benches some hundreds of metres high. Thus, at the southern extremity of the Taimyr depression a large fault (with an amplitude of the order of 1000 m or more) has been encountered by the author on the Ikon River (the right-hand side tributary of the Pyassina River). Here from the side of the lowlands Lower Cretaceous coal-bearing deposits nearly abut on the mountains of the Central Siberian plateau composed of Permo-Triassic tuff-lava beds (traps of the upper division of the Tunguska series). Mineralogical analyses of the latter have shown a complete absence of products of decomposition of traps, *e. g.* of pyroxene grains, which testifies to a very young (Post-Lower Cretaceous) age of the fault. The pitch of the Permo-Triassic to a great depth along the line of this fault is confirmed also by geophysical data, namely by a change of the strength of the magnetic field, when passing through the fault, by nearly 1500 gammas and a change of the values of gravitation by 21 milligalls (I. G. Zemskov). Along the northern border of the depression on the left bank in the lower course of the Pyassina a similar fault has been described by N. N. Mutafi. The amplitude of the fault along which the Permian coal-bearing beds come in contact with Permo-Triassic tuff-lava beds, exceeds 1000 m. Southward, according to the present author's observations, within the rocks of tuff-lava deposits there exists another fault parallel to the former, with an amplitude of no less than 100 m. Along the fracture of the fault in a narrow gorge flows the Pura River, a left-hand side tributary of the Pyassina. Further East, the existence of fractures along the southern edge of the Byranga Mts. was presumed by F. G. Markov for the region of Lake Taimyr and, finally, in 1943 was established by T. P. Kochetkov on the eastern coast of Taimyr, South of Pronchistchev Bay

All these data are not sufficient, however, for considering the Taimyr depression a vast graben. Along the southern margin of the depression, both West and East of the fault discovered by the author on the Ikon River, there probably takes place a gradual pitch of the Palaeozoic of the platform, including the Tunguska series, towards the depression. This is expressed both in a northward pitch of the axes of Hercynian folds, developed in the Norilsk region, and in a gentle northward dip of the beds of the Tunguska series and of the marine Palaeozoic composing the northern marginal part of the platform throughout the area, begin-

ning with the basin of the Dudypta and ending with the Kotuy and Anabar. It is essential to note that the magnetometric and gravimetric profiles crossing the southern edge of the depression along the Avam, Kheta, Kotuy and Anabar rivers, do not mark the boundary between the depression and the platform either, being evidence of the absence of large fractures. A gradual pitch of Palaeozoic rocks towards the depression has been recorded also in the southern part of the Taimyr folded zones, owing to which along the border of the depression Permo-Triassic and Permian series are most extensively developed, superseded to the North by older series.

In the light of these data it would seem correct to consider the Taimyr depression as a gently sloping trough, at places complicated by faults along the borders. The limiting age of the downwarping at the place of the depression is determined by the development on both sides of the latter of Permian coal-bearing and Permo-Triassic tuff-lava beds, displaying no increase in thicknesses and no apparent change of facies when passing to the depression. Inside the latter, rocks of the Tunguska series have been encountered in drill-holes of the Ust-Yenissei port and Kozhevnikov Bay at depths of over 1000 m. The upper age-limit of the downwarping can be determined from the Mesozoic sediments, at the base of which lie sediments of marine Lower Triassic. In their actual development they do not pass the boundary of the depression, while within the latter their total thickness attains 2000—3000 m. Consequently, the appearance of the Taimyr depression may be referred to the Triassic, more precisely, to the time directly following the formation of tuff-lava beds, essentially corresponding, as it would seem, to the lower part of the Lower Triassic. If it be so, it would be natural to see in the Taimyr depression a frontal depression of the Hercynian folded zone of the Taimyr, the shaping of which, from all the data available at present, ended at the beginning of the Triassic. The downwarping of the depression proceeded throughout the Mesozoic, being later on probably associated with the folding of the Upper Yana region.

The original downwarping at the place of the depression had outlines different from the present ones, comprising the adjacent areas of the Siberian platform and the Taimyr folded zone. This is evidenced by the mineralogical composition of the Mesozoic sediments of the depression, which are products of decomposition of arid rocks (probably of North Taimyr and partly of Western Siberia) and, with a few exceptions, almost entirely deprived of minerals of the trap province, in the first place of pyroxenes. In the Mesozoic the northern part of the Central Siberian plateau and of the southern marginal part of the Byranga Mts. composed mainly of traps, were subject to no erosion, but on the contrary, served as areas of accumulation. This is confirmed by the presence in the Quaternary deposits of the depression, even at the bench of the Central Siberian plateau and the Byranga Mts., along with a large amount of pyroxenes (up to 33 per cent of the 0.1—0.01 mm fraction) also of quartz and acid plagioclases, constituting, as a rule, over 50 per cent of the grains. This is also testified by the presence in the Quaternary sediments within the tableland (the Karaelak plateau) of products of decomposition of acid rocks—quartz (up to 50 per cent of the light fraction), zircon (up to 2 per cent of the heavy fraction), and others. These minerals were probably introduced into the Quaternary rocks as a result of erosion of Mesozoic formations that had previously overlain the present highlands adjacent to the depression. The latter were raised only in the Cenozoic, possibly even in the Quaternary, the elevation being in a number of cases accompanied by fractures, which told also upon the present relief.

The elevation of the mountain massifs along the borders of the depression continues until now—during the post-glacial epoch the Central Siberian tableland rose by no less than 25 m with respect to the depression (judging from the height of the lake terraces at the northern extremity of Lake Pyassina in the depression and South of the lake, on the slopes of the tableland at Norilsk). Comparing the marks attained by the post-glacial transgression in the western part of Byranga (100 m) and inside the depression (40—50 m), we must conclude that

the vertical displacement in the area of Byranga with regard to the Taimyr depression was still greater during the post-glacial time (of the order of 40—60 m).

The fact that the Taimyr depression and the adjacent portions of the platform and of the Taimyr folded zone have retained mobility until to-day, is reflected also in the distribution of gravity anomalies. The work carried out by I. G. Zemskov and B. N. Kashighin has shown with sufficient clearness a regional negative gravity anomaly (up to—65 milligalls), stretching parallel to the long axis of the depression and comprising its southern part, as well as the northern marginal part of the Siberian platform. This zone includes the areas where the Lower Palaeozoic is exposed at the surface (the Kotui River), and the areas where the roof of the Upper Palaeozoic lies at depths of more than 1000 m (Kozhevnikov Bay). In the Norilsk region the zone of negative values of gravity crosses gently sloping Hercynian folds of a meridional direction, being in no way dependent upon the arrangement of the structural element.

All this is evidence of the formation of the given anomaly as a result of density variations at great depths, rather than of the roughness of the roof of the Palaeozoic or pre-Cambrian. On account of the non-uniform densities of the section, these factors also play a certain part, being superimposed upon the general anomaly and producing local anomalies connected with the structure of the upper portions of the Earth's crust.

Pendulum determinations made in the northern marginal portion of the platform have shown that the synclines filled with traps which are rocks of a higher density, possess an excess of masses, while the anticlines in which the tuff-lava beds have been completely removed by erosion, show, on the contrary, a deficiency of mass. Probably the same is true of the depression, although, because of a greater pitch of traps in the synclinal downwarpings and their approach to the surface in the anticlines, the maxima may occasionally be shifted to the limbs, or even to the axial parts of the anticlines. Gravity distribution must also depend on the position of the leading channel-ways of the trap magma. Concerning the Norilsk region, V. K. Katulsky arrived at the conclusion that such channel-ways are located in synclinal downwarpings, being confined to fractures open downwards, augmenting the increase in gravity in structural troughs. As a result, in spite of the deficiency af mass common for the North of the platform, in the synclines positive values of gravity are occasionally recorded (the Norilsk and Karaelak brachysynclines).

On the whole, the zone of negative gravity anomalies here considered has probably been formed in the following way. As has been shown, the downwarping of the depression and the accumulation of sedimentary beds within it proceeded since the Mesozoic. Judging from the great thickness of Quaternary deposits (up to 200 m in the lower course of the Yenissei), the sinking of the depression is still in progress, while the northern marginal part of the Siberian platform is rising. This brings about a rarefaction of the subcrustal substance under the uplifted areas, and consequently the appearance of a negative anomaly of gravity. Being caused by changes in the state of matter at great depths, this anomaly cannot strictly coincide with the position of the rising blocks, and as a result of this affects also the southern border of the depression. Further North, in the central parts of the Taimyr depression, a prolonged sinking caused an increase in the density of abyssal masses and, as a consequence of this, the appearance of positive gravity anomalies (up to 50 milligalls), which is characteristic of the greater part of the area of the depression.

The local minima which are occasionally observed here, too, have probably been caused by the peculiar features in the structures of the upper strata of the Earth's crust, *e. g.* by the erosion of traps within definite areas. It is only at the northern border of the depression that negative gravity anomalies (up to—30 milligalls) become predominant again; these are probably associated with the elevation of the Byranga area. Within the Taimyr folded zone small positive anomalies still predominate, which are due to an increase in the density of matter as a result of Hercynian folding and of intrusions.

The tectonic activity of the Taimyr depression is confirmed also by the growth of salt domes within its boundaries, which seems to continue up to the present time. Until recently, salt domes rising in the shape of bald hills above the general plain surface of the depression, and having an age so young that they have not been as yet levelled up by erosion, have been known only at the eastern extremity of the depression, in the region of Nordwick. In 1942 the author discovered in the centre of the depression, in the upper course of the Dudypta River (71°32′ N. lat. and 94°10′ E. long.) an isolated bald mountain—the Chagdai, rising nearly 40 m above the surrounding late Quaternary alluvial terrace. The mountain is composed of Lower Cretaceous coal-bearing deposits, not exposed at any point in the adjacent regions, which makes one think of a tectonic, possibly salt-dome nature of the uplift.

In 1943 Mt. Chagdai was subjected to additional study by the expedition of the Norilsk Combinate, which established a periclinal attitude of Cretaceous sandstone beds on its slopes. Unfortunately, the possibility of the sandstones sliding down the slopes was not considered, so that it does not seem altogether impossible that the given mountain is an erosional monad rock. This is scarcely possible, however, since Mt. Chagdai is not unique. 100 km East, in the basin of the Boganida River, an analogous mountain, also with exposures in the Cretaceous coal-bearing series on the summit. was seen by A. F. Middendorf. In general, isolated mountains in the tundras of the Taimyr depression are of rather frequent occurrence, in most cases representing freezing knolls—«bulguniakhs» (their height, *e. g.* of the Yaryghin bulgunniakh, in the upper course of the Dudypta River, may attain 28 m). However, among the bald mountains, as shown by the above presented facts, there are mountains of a different, apparently tectonic origin, the discovery of which is very important for oil and salt prospecting, and should be the aim of the works to be carried out in the near future.

Received
7. VIII. 1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

GENETICS

A COMPOSITE PERENNIAL *ELYMUS*-WHEAT-*AGROPYRUM* HYBRID

By E. J. SOULIER

(Communicated by N. V. Zizin, Member of the Academy, 3. VIII. 1944)

In opening new vistas in plant breeding, N. V. Zizin wrote in 1935 «we will not dwell upon the enumeration of the intraspecific polymorphism of wild plants which could be employed for crossing definite cultivated plant; we shall instead try to show diagrammatically the ways of more remote intergeneric crosses which are possible from our point of view: 1) *Triticum* × *Agropyrum*, *Triticum* × ×*Elymus*; 2) *Secale*×*Agropyrum*, *Secale*×*Elymus*; 3) *Hordeum*×*Elymus*, *etc.*».

The method of remote hybridization which yielded brillant results in the intergeneric crosses wheat-*Agropyrum* is being successfully worked out by a group of workers under the immediate guidance of N. V. Zizin.

According to the above scheme, aside from the principal direction (*Triticum*× ×*Agropyrum*), the technique is being worked out and hybrids are obtained in the following divisions: *Secale*×*Agropyrum* (G. D. Lapchenko); *Triticum*×*Elymus* (V. E. Pissarev); *Hordeum*×*Crytesion* (N. M. Vinogradova); *Hordeum*×*Elymus* (F. Kh. Borkhteev).

We obtained in 1943 for the first time a composite perennial *Elymus*-wheat-*Agropyrum* hybrid.

It should be noted that thousands of crosses between wheats and the representatives of the genus *Elymus* which were attempted both at Omsk and in Moscow failed until recently.

Numerous attempts made abroad similarly remained fruitless.

Like any other new way in plant breeding, the crosses with *Elymus* are promising both from practical and theoretical point of view. The attractiveness of the biological resistance, high productivity and longevity of *Elymus* for the creation of perennial wheats is obvious.

For crosses with *Elymus* we have employed, aside from the typically annual cultivated plants, mainly the perennial forms of wheat-*Agropyrum* hybrids, One of such forms, which was a F_4 perennial sterile wheat-*Agropyrum* hybrid, pollinated with pollen of *Elymus giganteus* yielded in 1943 two hybrid kernels.

These kernels were germinated in February, 1944, under laboratory conditions. The seedlings which appeared from them were needle-shaped and slightly coloured with anthocyanin. The first shoots grew very slowly. The anthocyanin coloration at the lower side of the leaf blade and on the sheath of the first leaves covered with very short prickles persisted for a very long time. Cilia along the edge of the sheath were absent. The ligule was pellicular, white and gradually transgressing into the sheath. No auricles were present. At the base of the leaf blade there were 2—3 very short cilia.

The leaves were long, narrow, of pure green coloration, finely serrate at the edges, many times wound into a spiral. Venation was clearly expressed. Along

the veins there were very short prickles, owing to which a slight roughness of the leaves was noticeable.

The leaves on the main shoot were situated approximately opposite to each other; the secondary shoots (up to the fourth) formed at the leaf axils during tillering, were arranged in a common plain.

The plants were developing extremely slowly and not synchronously. During 3.5 months from February 15 to June 1 one of the plants attained the height of 6 cm measured from the soil to the base of the blade of the uppermost leaf. The main shoot of this plant developed seven leaves. The length of the uppermost leaf was 20 cm, and its width, 8 mm. At the axils of the lower leaves four shoots were formed with three leaves on each. The other plant was much behind the one just described, and had developed during the same period only the main shoot with five leaves.

At the period from June 1 to July 15 the first plant produced abundant tillers and formed spike-bearing stems within the culms. The shoot formation continued at several points, which is an indication of the creeping habit of the root system of the hybrid.

From the above description of the characters of the plants and from our observations on the course of their development it follows that the hybrids have inherited from *Elymus giganteus* the general habit, the rhythm of development, the creeping habit of the root system and other peculiarities of the leaf and stem structure.

A cytological study of the hybrids and of the maternal form, conducted at our laboratory under the guidance of V. A. Arnoldi, has shown that the somatic cells of the maternal form contained 48 chromosomes, while in those of the hybrid there were 52 chromosomes. As in the maternal form $n = 24$ and in the male, *Elymus giganteus*, $n = 14$, one could infer that an unreduced gamete of *Elymus giganteus* took part in fertilization.

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of the Non-Black Soil Belt.

Received
3. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

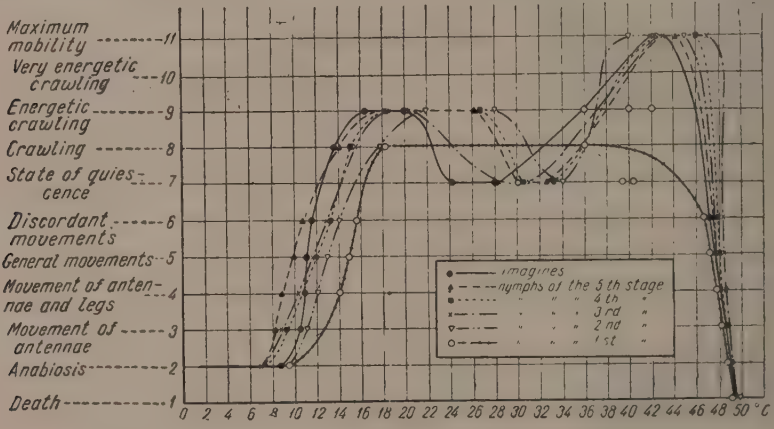
ECOLOGY

BEHAVIOUR OF *EURYGASTER INTEGRICEPS* PUT. AT DIFFERENT TEMPERATURES IN THE EXPERIMENT AND IN THE FIELD

By A. MACHOTIN

(Communicated by I. I. Schmalhausen, Member of the Academy, 6. VII. 1944)

The experiments and the observations here described were carried out in June 1941, near the town of Krasnodar *. In the laboratory experiments the insects were put into a glass containing a mobile thermocouple needle. The glass was put into a crystallizer, first filled with cryohydrate which reduced the temperature to—1°,—4°C. Cryohydrate was replaced first by cold, then by hot water, and the temperature of the glass was gradually increased to + 50— + 55° C. Throughout the experiments the humidity in the glass was maintained at 50 ± 3—4 per cent. Each experiment lasted from one to three hours. The activity of *Eurygaster integriceps* Put. was determined visually, no special calculations or measurement being made. The degree of the activity was recorded according to



a scale made up of eleven stages. The results of the experiments are brought together in the curves of the figure; along the abscissa is plotted the temperature, in °C: along the ordinate, the scale of activity.

With a gradual rise of temperature, the insects began to move at about 6—12°. Then the movements began to grow more rapid. At temperatures from 15 to 30° the activity attained its first maximum, without passing, however, the maximum

* The observations and the experiments were carried out together with N. S. Stsherbinovsky.

of energetic crawling. From 30 to 37° the insects remained in a state of quiescence. It may be noted that young nymphae responded somewhat later to the rise of temperature than adult bugs and nymphae of older stages.

A further rise of temperature caused a rapid increase in activity. Temperatures of 37—38° were marked by extreme activity, followed by decoordinated movements. At 49—49.5° death occurred in about 3—6 minutes. If the temperature did not cause the bugs any severe injury, their behaviour did not depend on the preceding temperature; *i. e.* a rise or fall of the temperature did not affect the behaviour corresponding to a definite temperature. The state of quiescence (24—34°) and of maximum mobility (40—48°) may be explained as temperature adaptations, in the former case protecting the insect from the excessive increase of body temperature; in the latter, helping it to avoid critical temperatures dangerous for its life.

The differences in the curve of the nymphae of stage I must be apparently explained by peculiarities of their behaviour under normal conditions. They do not feed, are almost inactive, and do not disperse unless touched or shaken from the plant upon which they are settled. At optimal temperatures they remain in a «state of quiescence». Their mobility is caused by the unusual conditions of the experiment. At high temperatures (from 36 up to 42°) the activity of some of the nymphae of stage I remains unaltered; it may either increase or decrease (see the symbols outside the curve for this stage). Thus, the temperature reaction is not stable in nymphae of stage I. The difference in the behaviour of young nymphae during the periods of critical high temperatures may probably be explained by the differences of the terms of stabilization of this reaction in different individuals.

The observations were made at wheat-fields on June 17, 18, 1941. A square plot, somewhat over 1 m^2 in area, of normal standing, was chosen in the middle of a wheat-field. The soil of the experimental plot consisted of chestnut czernozem of sufficient porosity, with many cracks and fissures. This plot was inhabited by several hundreds of adult bugs and by nymphae of different stages. The temperature of air and soil, humidity, *etc.*, were also recorded, along with observations on the behaviour of the bugs. The main results are brought together in the table. The minimum of humidity of the air during the 24 experimental hours was recorded at 3 p. m. (48 per cent); the maximum humidity, when dew was formed, between 4 and 5 a. m. (100 per cent). The maximum air temperature was recorded at 2—3 p. m. (34°), the minimum (17°), at 4—5 a. m. Temperature variations within the air layer at 10—50 cm above the soil were not great, ranging from 0.3 to 0.9°. Changes in the temperature of the soil surface in the shade repeat on the whole the changes of the air temperature. At 3 p. m. the temperature rose to 38.5—39°. Towards 4 a. m. it fell to 17—18°. The microclimatic differences of various soil plots are very great. These differences attain their maximum when the soil is under the action of direct sun rays. At 7 a. m. the temperature of the soil was 24—26°, differing from the temperature of the soil in the shade by 1—1.5°. At day-time the temperature of the soil attained 61.5° (2 p. m.). The temperature in the cracks and fissures of the soil remained comparatively stable, especially in fissures 15—20 cm deep, where it fluctuated within a range of 21.5° (4—5 a. m.) and 28° (4—5 p. m.). The temperature of the air recorded on June 17—18 can be considered almost equal to the average June temperature at Krasnodar, when there are scarcely cloud or wind.

Under the above-mentioned field conditions the bugs behaved in the following way. Before sunrise, all the insects of every stage remained at the top of the wheat stalks; after sunrise they crawled over to the east side of the field. At about 9—10 a. m. they crawled over to the shady side, and then to the ground. At about 11 a. m. not a single adult bug could be found on the wheat. Some were discovered in the soil fissures. At the same time, a horizontal migration of the bugs was observed. At about 2 p. m. the insects were observed to climb up the ears of wheat, and at about 4 p. m. they would crawl over to the sunny side of the ear. Many nymphae, especially of earlier stages, did not leave the wheat ears for 24 hours.

Hours	Humidity of air above the field, %	$t°$ of air above the field, °C	$t°$ of soil surface in the shade, °C	$t°$ of soil surface in the sun, °C	$t°$ of soil fissures at a depth of 15 cm, °C	Behaviour of *Eurygaster integriceps*
1 a. m.	96.5	18	20.5	—	23.5	Insects immobile at the top of wheat stalks
2 »	97.75	17.5	19.5	—	23	
3 »	98.75	17	18.75	—	22.5	
4 »	100	17	18	—	22	
5 »	100	17	18.5	—	22	After sunrise insects crawl over to the east side of the ear
6 »	92	21	21	—	22	
7 »	84	23	24	26	23	
8 »	81	26	25.6	30	23	
9 »	82	27	27.5	34	24	The bugs crawl over to the shady side of the ear and then to the ground
10 »	76	30.5	31.5	39	22	
11 »	83	34.5	35.5	43	23	Not a single adult bug can be found on the wheat. Some are discovered in soil fissures. At the same time horizontal migration observed
12 »	69	34	35.5	50	25.5	
1 p. m.	58	33	34.5	54	27.5	
2 »	53	34	35.5	61.5	24.5	The bugs climb up the ears of wheat
3 »	48	34	39	49.5	26.5	
4 »	50	33	34	44	28	The bugs crawl over to the sunny side of the ear
5 »	57	32.5	33	38	28	
6 »	72	28.5	30	—	26.75	
7 »	77	27.25	27	—	27	
8 »	87	22.5	25.25	—	26	
9 »	89	21	23.5	—	26	With sunset and gathering darkness the bugs' activity weakens. They crawl over to the top of wheat stalks
10 »	92	20	22.75	—	25.5	
11 »	95	18.5	22	—	25	
12 »	96	18	21	—	24	

Eurygaster integriceps feeds at any hour of the day, having no special feeding periods. The curve of the integument temperature of the bugs is smoother than that of the air temperature. This fact is accounted for by the insect's migration to places with a more favourable temperature, in other words, the behaviour of this bug has an adaptive character in relation to the surrounding temperature. In windless sunny days, the temperature is the main factor which determines the behaviour; and the behaviour in the field coincides with the behaviour in the laboratory under the same temperature conditions. The most susceptible to temperature variations are the adult bugs; the nymphae, especially of young stages, respond to temperature variations in a less conspicuous way.

In spite of the wide individual range of variation of the reactions to the temperature it is possible to establish mean indices which characterize the behaviour of the majority of bugs, and to indicate the limits of temperature for each state common to all the individuals. Thus, definite temperatures of the surroundings affect the majority of insects, determining their behaviour or position on the crop. Aside from temperature the behaviour of the bug is also affected by other factors of the surroundings, in particular illumination alternation (of day and night). *Eurygaster integriceps* should be classed among diurnal forms (being more active at the day-time).

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Received
6. VII. 1944.

Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

HYDROBIOLOGY

ALLO- AND AUTOCHTONIC REMAINS OF SUBFOSSIL DEPOSITS OF CONTINENTAL BASINS VIEWED UNDER THE ASPECT OF THE HISTORY OF THESE BASINS

By N. K. DECKSBACH

(Communicated by V. N. Sukachev, Member of the Academy, 23. VIII. 1944)

Among the analyses which serve to form a judgment as to the postglacial conditions, postglacial vegetation, and the history of the basins as a part of the landscape, an important place is due to the pollen analysis. Pollen analysis was suggested by Post as far back as in 1916, since when it has enjoyed a wide repute and universal application.

The analysis of other remains of subfossil deposits of continental basins (both of allo- and autochtonic origin) is much less widespread, and has not been developed to a degree of sufficient perfection. The group of remains belonging to the diatoms has been somewhat more fortunate in this respect (diatomic analysis).

It is obvious, however, that the analysis of pollen, as well as the analysis of any other fragment of vegetable or animal remains, inclusive of diatoms from subfossil deposits of basins, are supplementing one another, and that the analysis of diatoms may be helpful in a more thoroughgoing investigation of palaeobasins.

Unfortunately, at the present time we are lacking any generalization or critical review of the material bearing upon the analysis of animal and vegetable remains found in the subfossil deposits. In what follows the author makes the first attempt to fill up this lacuna.

When analysing animal and vegetable remains of subfossil deposits of continental basins, the explorer has largely to deal with autochtonic material; unlike this, pollen analysis is largely concerned with allochtonic material, genetically alien to the given basin, and which might occasionally have been brought to the basin from rather remote regions. Thus, for example, the pollen of the pine-tree may be transported over a distance of 700—900 km (4).

The results of analysis of autochtonic material are, of course, no less valuable, when forming an idea of the history of a given basin, than results based upon allochtonic material brought from adjoining regions, or especially from remote ones, on account of a number of specific characters peculiar to any given basin which have to be taken into account. Aside from macroclimate, conditions of microclimate become operative in this case, and a basin as a whole may have a definite influence even upon the climate of the surrounding region. Microclimate is to be taken into account because continental basins pass definite development stages before they attain their actual state. Thus, for example, definite morphometrical characters of a basin and an abundance of organic nutrient matter in it bring about the stage of prevalence of *Cyanophyceae*. This period in the develop-

ment of a basin is, among other peculiar features, characterized by that during this time the basin is absolutely independent of air temperature, and during summer its water is heated to a higher temperature than before or after this stage.

Analyses of pollen and of other subfossil deposits may occasionally yield discordant results. From the standpoint of the representatives of the «pollen analysis» trend in science, subfossil mud deposits of basins over vast areas of continents have been formed simultaneously; whereas according to the conceptions of limnologists, these deposits have been accumulated in different basins, occasionally located close to one another (lakes), at a different time, according to a number of conditions, especially to the morphometry of a given basin, to the degree of the development of organic life in it, *etc.* The depressions of different lakes have been formed and are formed at a different rate; therefore, forming an idea of the past of a given basin, the results of both kinds of analyses have to be taken into account.

The actual state of the analysis of animal and vegetable subfossil remains is characterized by the following main features.

I. A fragmentary nature of the records*.

Remains of definite animal species which play an important part in the life of basins, as the *Roratoria*, are practically missing**. Missing are also such importants groups of lower crustaceans as *Copepoda*, certain shell-less *Protozoa* ***, the *Cyanophyceae*, *etc.* Yet the remains available, in particular, the remains of organisms-indicators (see below), yield a material sufficient to form an idea both of the history of a basin and of the development stages through which it passed.

II. The existing subfossil records have not been worked out in due detail; the main rules how to make use of the data they contain have been clearly outlined, however.

a) The investigator engaged in the study of remains must not content himself by determining the species of the material discovered by him, but must attempt to a further determination of lower taxonomic categories, *viz.* varieties, morphs, *etc.* This will enable him to develop the analysis of the past of a given basin in more detail.

Among animal remains, wide prospects are opened by the study of the remains of one of the predominant groups of plankton, the *Cladocera*; among the benthos, the group of *Rotatoria*; the same holds true of the sponges. Among vegetable remains let us cite the diatoms.

b) In connexion with the attempts made to found the palaeoecology of subfossil forms, one may start from the conception that the ecology of many a recent form has changed but little if at all as compared to the ecology of their ancestry that had lived during the glacial period. Among algae this is true of such a leading group as the diatoms ([2],[6]). This fact enables us to interpret both certain palaeoecological features of organisms and many features of the past of the basins, restoring a link between the past and the present, and *vice versa*.

III. Among the remains of subfossil deposits special attention is due to the so-called organisms-indicators, which are very important when forming an idea of the history of a given basin and of the separate stages of its development.

The remains of organisms-indicators make it possible to judge.

1. Of certain chemical properties of palaeobasins, of the salt composition of these basins. The following cases should be distinguished:

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* The same is true of the method of pollen analysis and palaeontology, as a whole, which does by no means prevent these methods from developing in a most successful way.

** In the special atlas of remains of animal organisms, compiled by Rossolimo ([11]), the *Rotatoria* have not been mentioned at all. We met them, though in small quantities, in the very upper layers of silt and mud of basins of medium depth (Beloye Ozero, Kosino, Kliasma, Moscow-Volga channel), and in shallow basins (Chernobrovskoe lake, Sverdlovsk province, *etc.*).

*** According to our data, the *Rotatoria* are present in small numbers in the silts of various lakes of Sverdlovsk and Cheliabinsk provinces and in subfossil sediments. It is premature, however, to speak in this junction of a regional peculiarity of the Urals.

a) A basin that was poor in calcium, as testified by abundant remains of calciophobe forms of diatoms representatives of the genera *Eunotia*, *Pinnularia*, *etc.*).

b) A basin was rich in calcium (abundance of remains of molluscs). The fact that the basin had at a given time been at a stage of calcium abundance is testified also, aside from the abundance of molluscs, by a poor development of other organisms, in particular, of bed organisms. The deposition of calcium was in this case of chemical rather than biological nature.

c) A basin was rich in silicon, while the calcium content was a low one, and the content in soluble phosphorous compounds was high. Certain amounts of free aggressive carbon dioxide were present (abundant remains of diatoms, formation of layers of diatomite 6.5—7.0 m thick).

d) A basin was rich in calcium and at the same time, in humic substances. The existence of soft humic waters is evidenced by remains of sponges *Spongila arctica*, which are forms peculiar to such waters.

2. Of certain physical properties of the basin. These are: the high transparency of water with a low clay content (a feature peculiar to diatomic basins).

3. Of the silt contents of the basin. Abundant remains of siliceous spicules of sponges testify to a low degree of a given basin in the past; a high silt content of shallow lakes or pools prevents an abundant development of sponges.

4. Of the thermic and perhaps also morphometrical features of the basin. Presence of important numbers of morphs of definite organisms, for example, of lower crustaceans, the *Cladocera*, characterizing definite seasons: summer, autumn, winter, or spring. Sometimes a single morph is discovered in a basin, for example the hibernal one. Examples: different seasonal morphs of *Bosmina*, *Chydoridae*, *e. g.* ,*Acroperus*, *etc.* The presence among the remains of thermophilic forms or morphs, on account of the conditions of microclimate, landscape, or microclimate and local conditions.

5. Of the stability or, on the contrary, variability of the fundamental hydrological regime of the basin as a whole or of its separate parts in the past, as for instance of bays. Prevalence of definite genera of diatoms throughout a period corresponding to the formation of a nearly seven-metre-thick layer of diatomite in the basins of the Kola Peninsula is evidence of a stability of the main physical-chemical regime of the basin, and of the fact that its state did not differ from the recent one. This stability of conditions does not preclude, however, certain fluctuations in the conditions of the medium under which the diatoms were developing. According to the stability of the palaeoecological conditions of a basin, it is possible to conclude as to the stability of the conditions that prevailed in the surrounding regions during the same period of time.

6. Of the development of beds, the degree of development of the littoral zone and the nature of the basin as a whole (shallow basins). Presence of both qualitatively and quantitatively abundant remains of lower crustaceans (various *Chydoridae*), of remains of insect larvae (*e. g. Chyronomidae*), bottom bed diatoms, *etc.* at most different points of the same horizon.

7. Of the presence of affluents, rivers or streams feeding the basin. A diatomite contaminated with mineral elements makes one assume the existence of tributaries at the period of the formation of the diatomite.

8. Of the phenomena of important transformations and conversions of substances suspended or dissolved in the water. Occurrence of abundant remains of *Cladocera*, typical representatives of the group of sedimentors, permits of the assumption that the palaeodeposits underwent important transformations when still in the process of formation.

9. Finally, of certain fundamental morphometrical features of lake baths. Abundant resources of diatomites characterize basins provided with rather numerous bays well protected from currents, with a highly dissected shore-line, long spits, *etc.* A basin is either closed, or poorly running.

The morphometry of a palaeobasin may also be estimated according to the mode of distribution of the remains found in it. A uniform distribution of the

remains makes one conclude as to the fact that no transfer of material from the shallower, riverain parts of the basin to the deeper one took place at all, or else that such a transfer proceeded at a very low rate, because of the shallowness of the basin. Along with this, it is also possible to assume that the basin was well protected against the action of winds.

A further study of the organisms-indicators will, without any doubt, supply a number of valuable data which will greatly contribute to our knowledge of the past of basins and of the course of their development.

Received
23. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

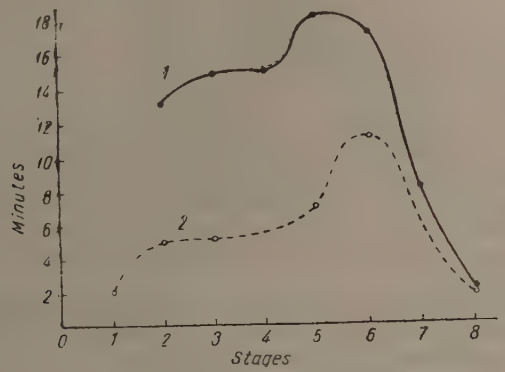
PLANT PHYSIOLOGY

A CONTRIBUTION TO THE STUDY OF HEAT RESISTANCE OF THE LEAVES OF COTTON GROWN ON SALINE SOILS

By A. KLESHNIN

(Communicated by A. N. Each, Member of the Academy, 20. IX. 1944)

The problem of heat resistance of the leaves of the cotton plant is of both practical and theoretical importance. Its practical bearing is clear from the fact that cotton is cultivated mainly in the republics of Central Asia where summer temperature may attain considerable value (above 40°), and salt accumulation in the soil is of a rather common occurrence. It should be anticipated *a priori* that salt ions which penetrate into the plants' cell would reduce the charge of the particles of the plasm and cause its coagulation at a lower temperature, thus decreasing the heat resistance of the leaves. This has been proved by Kaho (1) for a number of anions and cations. The experimental results of the present study are also in agreement with the above statement.



Heat resistance of the leaves of the cotton 8517. Water cultures. Sowing June, 4. Addition of salt (NaCl): beginning by June 6—0.1 per cent; beginning by June 21—0.2 per cent; beginning by July 3—0.3 per cent. Heat resistance determination, August, 31. *1*, control; *2*, salt accumulation

The heat resistance of the leaves of cotton was determined after Mazkov (2). The leaves belonging to a definite internode were cut off and placed into a beaker containing water; the temperature of water was gradually raised to 55°C. The leaves were kept at 55—60°C during ten minutes, after which they were transferred to cuvettes containing cold water; the water was then replaced by an 0.2 *N* solution of HCl. The appearance of yellow spots under the influence of HCl was

considered as an index of heat resistance; the earlier they appeared, the less was the resistance of the leaves tested.

The material of the first experiment were 57-day-old cotton plants of the variety 8517, which had developed in a water culture on full Hellrigel's nutrient solution. The salts were added in the form of NaCl (0.1—0.3. per cent of Cl). The results of the experiment are presented in the accompanying figure. It may be seen from the curves that cultivation under the condition of salt accumulation results in a strong reduction of heat resistance in cotton plant, the effect being stronger in the lower leaves. The last circumstance is evidently connected with an uneven distribution of the NaCl ions within the plant.

Analogous results were obtained from an experiment where the leaves of the 74-day-old cotton plant, variety C 460, which had developed under normal conditions (no excess of salt) were placed with their petioles for three hours into 2 per cent van't Hoff's solution (Table 1).

Table 1

Heat Resistance of Leaves of Cotton C 460
(Field experiments; sowing, June 4; determination of heat resistance, July 20)

Internodes	Appearance of yellow spots under influence of 0.2 *N* HCl (min.)		Internodes	Appearance of yellow spots under influence of 0.2 *N* HCl (min.)	
	Control — leaves on water (3 hrs)	Exper. — leaves on 2 % van't Hoff's solution (3 hrs)		Control — leaves on water (3 hrs)	Exper. — leaves on 2 % van't Hoff's solution (3 hrs)
1	4	—	10	19	10
2	8	5	11	17	10
3	7	8	12	17	8
4	8	8	13	6	9
5	17	12	14	4	5
6	17	12	15	2	5
7	14	14	16	—	4
8	16	12	17	—	2
9	22	14			

Van't Hoff's solution was taken as a factor of salt accumulation on account of Henkel's ([3], [4]) observation that when seeds are treated before sowing with this solution, the salt resistance of wheat and cotton is greatly increased. A three-hour-long treatment with saline was sufficient to produce a considerable reduction in heat resistance of the leaves in the cotton. This was especially apparent in the middle leaves (of the 5th to the 12th internode) which are physiologically the most active ones.

The experiments with the solutions of NaCl of different concentrations, into which the leaves of cotton 8517 were dipped with their petioles for 2.5 hours, have shown that the decrease in heat resistance went parallel to the concentration of salts in the medium (Table 2).

Table 2

Heat Resistance of the 10th Leave of Cotton 8517 Treated with NaCl Solution by Dipping its Petiole for 2.5 hours
(Field experiments; sowing, June 4; determination, July 20)

Concentration of NaCl expressed in % of Cl	0	1	2	3	4
Time of appearance of yellow spots under influence of 0.2 *N* solution of HCl (minutes)	18	13	10	9	4

Heat resistance of leaves belonging to different internodes is of considerable interest. The variation of the heat resistance of the leaves with internodes is expressed by Krenke's curve ([5],[6]); the maximum resistance had place in the middle leaves, while minimum resistance is found in the lowest leaves whose general age is less, and in the uppermost leaves whose proper age is insignificant (see figure).

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PLANT PHYSIOLOGY

RELATION BETWEEN SEEDS AND PERICARP IN THE PROCESS OF FRUIT GROWTH AND RIPENING

By J. V. RAKITIN

(Communicated by A. N. Bach, Member of the Academy, 26. IX. 1944)

A rather large number of studies have been devoted to the growth and ripening of fleshy fruits, such as, for instance, apples, oranges, tomatoes or peaches. They suffer, however, from a serious drawback, namely no due attention was paid in them to the problem of the interrelations between seeds and pericarp, which is not only of theoretical but also of great practical importance. One should not leave out of consideration, first, that the ripening of pericarp begins at the period when the seeds become mature. Until the moment when the embryos are formed within the seeds, the reserve materials stored and the outer covers developed, the tissues of the pericarp remain hard, have unpleasant taste and green coloration. As soon as the seeds have attained maturity, sharp changes occur in the pericarp: its tissues are softened, the unpleasant taste disappears. Soluble carbohydrates are accumulated, there appears odour, chlorophyll is destroyed, and the surface of the pericarp, sometimes all its tissues become intensely coloured.

In the plants that form fleshy fruits, the seeds are distributed by animals which feed on the fruits; the adaptive nature of the relations between the seeds and the pericarp can consequently be interpreted in the following way: until the seeds are ripe, the pericarp allures the animals neither by its coloration nor taste, this preventing the seeds from being destroyed. As soon as the seeds have ripened and formed a protective coat, permitting their passage through the digestive tract of the animal without being destroyed, the pericarp begins to ripen and to attract animals.

From the studies by Müller-Thurgau (1) and by others (2) it is well known that apples, pears, dates, gooseberries and fruits of many other plants are the larger, the more seeds they contain. It has been proved by Ewert (3,4) that when parthenocarpic fruits are present together with seed-bearing ones, the latter develop far better than the former. No less illustrative is the influence of the seeds upon the shape of the pericarp. In asymmetric fruits it is usually the part containing more seeds which develops better.

The effect of seeds upon the pericarp is especially clear when the rate of maturation of parthenocarpic fruits is compared with that of fruits containing seeds. It has been shown, for example (2), that in the Japanese persimmon *Zengi* the seed-bearing fruits usually mature toward the end of July, while the parthenocarpic ones ripen towards December only.

It is of special interest that changes typical of the maturation of the pericarp appear in many fruits first in the tissues which are nearest to the seeds, and only later in more remote regions of the pericarp. This regularity was traced by us

in the fruits of tomatoes, melons, cucumbers and in others. The manifestations of this regularity are illustrated by data contained in the table.

It follows from the table that the formation of mucous in the tissues surrounding the seeds, which is the first index of ripening of pericarp, does not begin until the seeds have attained physiological maturity.

Course of Various Processes in Maturing «Humbert»-Tomatoes

Coloration of fruits	Condition of tissues surrounding the seeds	Seed maturity	Place of sampling	Acetaldehyde content, mgs per 100 gm of tissue	Peroxydase activity, cm^3 of 0.1 *N* permanganate
Green	Hard	Immature	Periphery	0.11	6.4
			Middle	0.13	6.9
Pale-green	Traces of mucous	Mature	Periphery	0.12	5.0
			Middle	0.22	1.2
Yellow-green	Fully mucilaginous	»	Periphery	0.18	4.3
			Middle	0.32	0.8
Pink	Do	»	Periphery	0.26	2.8
			Middle	0.36	0.5
Red	»	»	Periphery	0.31	1.2
			Middle	0.40	0.3

Along with the development of mucous there occurs in the tissues of pericarp accumulation of acetaldehyde and a drop in peroxydase activity; these changes are manifest earlier and more intensely in the tissues adjoining the seeds than in peripheral ones. Determinations of acetaldehyde were made after the method described by us ([5]); those of peroxydase, after the method by Sukhorukov *et al.*([6]). Since these alterations, as well as the process of mucous formation, are the indices of pericarp maturity, it follows that maturation processes first appear in the tissues adjoining seeds and only later in peripheral ones.

All these facts indicate that the impulse causing the maturation of the pericarp of the seedy fruits, is conveyed from the seeds. This is supported by the fact that the flesh of the seedless fruit, in distinction to the seedy ones, matures evenly throughout its mass. The fruits of the Japanese persimmon may serve as a very instructive example ([7]).

Finally, it should be noted, in order to characterize the relations between the seeds and the pericarp, that the seeds do usually not germinate within the juicy fruits, while when taken out of the fruit, they germinate rather quickly. This phenomenon should probably be regarded as one of the adaptations to the distribution of seeds by animals. Had the pericarp not preserved the seeds from germinating, no distribution by fruit devouring animals would be possible, since germinated seed would, of course, be destroyed. The seeds germinate within the pericarp only if the latter is over-matured and cannot serve as a good lure to animals.

Starting from the fundamental fact, *viz.* that the maturation of the pericarp does not begin until the seeds are mature, one can easily understand that there exist two distinct periods in the interrelation of the seeds and pericarp: the first period comprises the time from the moment of fertilization to seed maturation; the second, from the seed maturity to full maturation of the pericarp. The numerous and well-known results obtained in the chemistry of the development of the fleshy fruits make it possible to assume that these stages are clearly distinct too, from the point of view of metabolism (synthesis $\rightleftarrows$breakdown).

During the first period in the seeds and pericarp processes of synthesis predominate, which are connected with cell growth, with the increase in the mass of the living matter and with the storing of high-molecular compounds as reserve materials. During the second period there appear essential differences between

the metabolism of the seeds and of the pericarp. The second period is characterized in the seeds by a general decline in their biochemical activity, and by a transition to a state of «rest», while in the pericarp there occurs a decrease in synthesis and an increase in hydrolytic breakdown.

While noting that the processes of breakdown predominate in the pericarp at the second stage of fruit development, it is necessary to mention that certain synthetic processes, *e. g.* formation of essential oils and of pigments, take place at this time at a higher rate. These processes take, however, no important part in the general metabolism. If the fruits are not destroyed in some way when they reach maturity, the processes typical of the second period will ultimately lead to overmaturing and to the death of the cells of the pericarp. The possibility of a spontaneous destruction of the pericarp, *i.e.* without participation of microorganisms, has already been demonstrated by Brefeld ([8]). Ultimate breakdown of the pericarp, followed by the liberation of the seeds, usually occurs under cooperation of microorganisms.

The natural end of life of the fruit as a whole is the complete breakdown of the pericarp and the liberation of the seeds, giving rise to a new plant generation.

Seedless fruits are very rare exceptions in wild plants, while in a number of cultivated plants, such as banana, apple, pear, mandarine, grape, gooseberry, cucumbers, it is of a rather wide occurrence and should be regarded as a result of long cultivation. In contrast to seedy fruits, parthenocarpic ones have, as mentioned above, a number of peculiarities; however, as regards the basic exchange (synthesis$\rightleftarrows$breakdown), they are characterized by the same two periods, as in the case of seedy fruits.

During the first period, embracing the time of cell growth and storing of high molecular substances, the processes of synthesis predominate, while during the second period corresponding to the time of maturation, there is a predominance of the processes of hydrolytic breakdown.

All these observations make it possible to believe that the fundamental regularities, typical of the development of seedy fruits, are the ultimate expression of the interrelation existing between seeds and pericarp, and that a correct understanding of the peculiarities of the growth and maturation of fruits arising parthenocarpically is not possible unless we have a clear idea of the regularities peculiar to seed bearing fruits.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PLANT PHYSIOLOGY

ON THE GROWTH HORMONES IN A DISEASED PLANT

By K. SUKHORUKOV and B. STROGONOV

(Communicated by A. A. Richter, Member of the Academy, 29. VIII. 1944)

Many diseases in higher plants are accompanied with considerable abnormalities in the growth processes and in morphogenesis, leading to tumours, warts, variation in size and shape of organs, appearance of new parts and to other changes which are widely known in phytopathology. At the present state of phytoendocrinology it is beyond any doubt that hormones take part in the growth, morphogenesis and development of plants ([4]). It is quite natural to suppose that in a diseased plant, too, the external changes and the internal conditions are connected with the hormonal changes. The study of plants affected by infectious diseases is hindered by important circumstances, *viz.* by the production of hormones by the causative organism, by the reaction of the two organisms to foreign hormones, and by the changes taking place in the hormones of the dead cells. An attempt is made in the present paper to discuss these complex interrelations.

Growth Hormones in Diseased Plants, in Secretions of Parasites and Saprophytes. The auxines weer determined according to bend of decapitated coleoptiles of oats after the method described by Čajlachjan ([5]). The hormone was extracted with sulphuric ether freed from peroxides, either from raw or from quickly exsiccated materials (at 60—80°C); the medium on which the parasites had been grown was also extracted. In preparing agar and agar-blocks we have used the method described by Boysen-Jensen ([1]). The hormones of cell division were determined by the yeast-method. Let us pass to the results.

The leaves of oats (*Avena sativa*), severely infected by *Puccinia coronifera* and healthy ones, were quickly dried at 80°C, the auxine was extracted from 4 g of material; the alleged auxines were mixed with 1 cm³ of agar. For the tests twenty coleoptiles were taken, the bends after 4 hours were printed on photographic paper and measured. To check chemically, nitrogen was determined in the ether extract. The results are given in Table 1.

Table 1

	Bend in degrees (average from 20)	Nitrogen content in ether extract in mgs per 100 g of leaves
Healthy leaves	2.2	34
Diseased »	1.4	23
Pure agar	1.4	—

No auxine was found in the diseased leaves.

Tubers of potato (*Solanum tuberosum*) variety Early Rose (highly susceptible) and Seedling 8670 (highly resistant) were sliced, and a fraction of the slices was inoculated with *Phytophthora infestans*; the rest were left for control without inoculation. After the mycelium of the fungus had abundantly developed, both the infected and control slices were ground to pulp under addition of a few drops of a 5 per cent citric acid solution. The auxine was extracted from 100 g of the pulp in each case. Determinations made as above yielded the following results:

		Bend in Degrees (average from 20)
Early Rose	healthy (control)	1.0
» »	infected	4.4
Seedling 8670	healthy (control)	2.3
» »	infected	1.8

In the infected susceptible variety there is an accumulation of auxines, while in the resistant variety the auxines are diminished. Seedling 6870, aside from its resistance to the fungus, is characterized by a high activity of the oxidases.

Bacillus tumefaciens is a causative organism of tumours with a peculiar hyperplasia after Naumov ([3]). Küster ([9]), along with many observations of various students and of his own, reports on the secreta of *B. tumefaciens*: Metchnikoff([2]) and Caspari ([7]) discuss plant warts as compared to animal tumours; Braun ([6]) reports interesting observations according to which killing bacteria by heat does not stop the autonomous growth of the tumour. We had a virulent strain of *B. tumefaciens* obtained from Prof. Israilsky. The bacilli were grown on a sterilized decoction of the seeds of white beans (1 : 10) during twenty days. Into some of the bottles containing *B. tumefaciens* tyrosin was introduced, 25 mg per 50 cm³ of the nutrient medium. After a twenty-day cultivation the contents of the bottles were filtered: the filtrates were acidified with citric acid and extracted with ether. The auxines from 100 cm³ of the nutrient medium were transferred to 1 cm³ of agar, and the medium was used, upon extraction with ether, for bios determination (1 cm³ to 20 cm³ of medium after Gajduk). In the ether extract nitrogen was determined, and qualitative reactions for indol made (with paradimethyl-amino-benzaldehyde). The results are presented in Table 2.

Table 2

Nutrient medium	Auxines, degree of bend of oat coleoptiles	Bios, mgs of dry yeast per 1 cm³ of medium	Ether-soluble nitrogenous substances, mgs per 100 cm³ of medium	Reaction for indol
Medium without inoculation (control)	1.40	0.6 (100,0)	1.49	none
Medium after growth of *B. tumefaciens*	2.3	4.6 (766.7)	1.49	»
Medium added tyrosin after growth of *B. tumefaciens*	5.9	5.8 (966.7)	2.61	positive

With saprophytic nutrition *Bacillus tumefaciens* is able to form auxines and hormones of cell division When tyrosin is included into the nutrient media, the content of ether-soluble nitrogen is increased by 75.1 per cent; there appears indol, and auxine production is augmented. In our opinion, the 50 per cent increase of bios in the tyrosin medium may be connected with the increased permeability of the cells of the bacilli to the bios in the presence of the auxine. *Bacillus tumefaciens* is clearly able to form hormones of the two types, the conditions of nutrition having a noticeable effect upon this process. Abnormalities in the growth

of the diseased plants may be explained by an excessive supply of hormones to the tissues affected.

Aspergillus niger is a typical saprophyte and, according to Boysen-Jensen, is able to form heteroauxine. We have cultivated *A. niger* on the glucose-mineral medium and on the same medium with a half of ammonium nitrate replaced by tyrosin. After 15 and 53 days of growth of the fungus the medium was analysed. The results are given in Table 3.

Table 3

	15-day-old culture (auxine measured in degrees of bend of coleoptiles)	53-day-old culture (derivatives of indol in 100 cm^3 of medium, after Imabuchi [8], 1 cm^3 $\frac{1}{400}$ $KMnO_4$)
Medium without tyrosin	4.2	27.4
Medium containing tyrosin	4.0	11.3

Aspergillus niger is capable of producing auxine, but the ways of its formation within the organism differ from the case of *Bacillus tumefaciens*. The latter shows a clear positive reaction for tyrosin, and the benzene rings are evidently readily converted into indol rings. Here the tyrosin is assimilated by the organism without a direct and rapid convertion into indol. In *B. tumefaciens* the ability to auxine production is a biologically important adaptation to parasitism.

On the *post mortem* Formation of Auxine. Many plant diseases are connected with tissue necroses. In the case of necroses dark areas arise, due to oxidative condensation of certain substances. After Raper and Evans [10] when tyrosin is oxidized by tyrosinase, the first product of oxidation, dihydro-oxyphenyl-alanine, gives in the course of subsequent oxidation compounds containing a condensed nucleus, derivatives of indol and undetermined substances, the melanines. Preparations of the tyrosinase of vegetable origin often contain other oxidases which affect the first stages of oxidation and decrease the formation of dihydro-oxyphenyl-alanine. It was of interest to clear up the possible *post mortem* syntheses of auxine which may contribute to the knowledge of the protective reactions of the plant, expressed in oxidative necroses.

In solving this problem use was made of potato tubers (variety Woltmann) which produce oxidative necroses in the case of many diseases and contain active tyrosinase. The tubers were ground into a homogeneous pulp to which were added various substances susceptible to oxidation and capable of taking part in the reactions of condensation. The potato pulp was kept twenty-four hours under the vapour of chloroform, and the extracts of auxine were prepared of it. For each test 100 g of the pulp was taken. The results of several determinations are presented in Table 4.

It can be inferred from this table that no formation of plysiologically active hormone takes place in the course of oxidation of the various subtances and in the presence of substances which are capable of reaction. On the contrary, the auxines which survive the death of the tissue, disappear later on. No explanation is offered for the latter phenomenon. Judging by the ether-soluble nitrogenous substances, no indol derivatives are formed either (the determinations of ether-soluble nitrogenous substances were, of course, of rather provisory character). The results obtained, as regards the vegetable enzymes, do not contradict Raper's ideas [10], but do rather confirm them; they are, moreover, in full accord with the earlier data on the considerable decrease of auxines in the diseased plants. There takes place evidently breakdown of the auxines in the dead cells with protective necroses.

Table 4

Material studied	Bends of coleoptiles, in degrees	Ether-soluble nitrogenous substances, mgs of N per 100 g
1. Potato pulp	2.8	0.00
» » plus 0.5 g of tyrosin	1.6	0.00
2. Potato pulp	1.8	0.00
» » plus 0.5 g of tyrosin	0.8	0.00
» » plus 0.5 g of tyrosin plus 0.14 g of acetaldehyde	0.9	0.00
3. Potato pulp	3.3	0.00
» » plus 0.21 g of acetaldehyde	4.4	0.00
4. Potato pulp	3.6	0.00
» » plus 0.5 g of hydrochinone plus 0.35 g of glycocol	0.8	0.00
» » plus hydrochinone plus 0.35 g of glycocol plus 0.22 g of acetaldehyde	0.7	0.00
5. Potato pulp	4.1	0.00
» » plus 0.5 g of pyrogallol plus 0.3 g of glycocol	2.8	0.56
» » plus 0.5 g of triptophane	3.0	0.00

Note. The potato pulp, both pure and under addition of various substances, prior to extraction was rapidly dried in a stream of heated air.

Let us make some general conclusions. In diseased plants the auxines are destroyed in the case of protective necroses, while no auxines are formed in dead cells. Parasitic organisms which disturb the correlation between the growth processes (*B. tumefaciens*) are capable to produce and secrete auxines and hormones of cell division.

Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

PHYTOPATHOLOGY

TYROSINASE ENZYME IN THE CAUSATIVE AGENT OF THE BLACK BACTERIOSIS OF CEREALS

By A. P. KLYKOV

(Communicated by A. A. Richter, Member of the Academy, 19. IX. 1944)

Ervin Smith (1) published in 1917 an article in which he was describing a new disease of wheat, for which he had coined the name of «black chaff». This name has soon become current. According to Smith, the causative agent of the disease was *Bacterium translucens* var. *undulosum*.

In the subsequent, number of research workers (3,4) became interested in the biochemical processes taking place in tissues under the influence of the life activity of bacteria; no satisfactory explanation could be given to this phenomenon, however. In the present author's opinion, the phenomenon might be due to the effect of the oxidizing enzyme, tyrosinase, upon the product of hydrolytic protein splitting, the tyrosin, which is present in plants in large quantities. Now it was necessary to demonstrate that the tyrosinase enzyme was present in the causative agent of black chaff.

With this aim in view there was prepared a special agar medium to which 0.5 per cent of tyrosin was added. This medium was sown with 20 strains of *Bact. translucens*.

After several days the medium assumed the most various shades of black, from the deepest black and to normal yellow. Sowing was repeated and gave analogous results.

This experiment went to confirm the interpretation given by the present author to the blackening of tissues; it has shown, moreover, that different strains of the same bacterial species do not contain equal quantities of the tyrosinase enzyme, and that there are cases where the enzyme is absent altogether.

Received
21. VIII. 1944.

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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

ZOOLOGY

DISTRIBUTION OF BLACK HAMSTER IN THE UKRAINIAN SSR

By S. GERSHENSON

(Communicated by I. I. Schmalhausen, Member of the Academy, 3. VII. 1944)

The most promising way for the investigation of the laws of natural selection is the study of the quantitative changes which take place in wild populations of a species in the course of several generations. The hamster is a very suitable object for such investigations. It has been know long since that wild populations of hamsters often contain a high percentage of melanic individuals. As shown before ([1],[2]), the difference between common and black hamsters is genetically very simple—black colour is inherited as a simple Mendelian dominant, and in wild populations black and common individuals mate at random. It is known that the percentage of melanists greatly varies from population to population. Lastly, it is very important that hamsters are the object of a rather developed branch of fur-trade, and we had the opportunity to obtain from numerous factories of «Soiuzpushnina»* data concerning the number of black and common hamsters, trapped there during a number of years.

The present work is based on the records of 529 district-factories of «Soiuzpushnina», distributed all over the Ukrainian SSR, except its western regions. The records include data concerning about 2 millions hamsters trapped there during 1934—1939. As material for the present work served also the results of extensive field collections gathered by the author during 1939—1941.

An analysis of the trapping methods and of the technique of the recording of hamster-pelts purchased by the local factories leads to the conclusion that the data of «Soiuzpushnina» reflect very exactly the actual frequency of melanists in the corresponding districts. This is fully confirmed by a direct test made by large-scale field investigations in several localities.

In Table 1 are given summarized data concerning the total number of hamsters and the number of black hamsters trapped during 1934—1939 in different regions of the Ukrainian SSR.

As can be seen from Table 1, the frequency of black hamsters greatly differs in different regions of the territory investigated. Our data show further that such differences exist in different districts within each region. At the same time the frequency of melanists in each district remains relatively constant throughout the period studied, *i. e.* the distribution of melanists varies but insignificantly from year to year.

Black hamsters occur on a large part of the investigated territory, though the area where they are met is considerably smaller than the area occupied by the species. Within the boundaries of the area of distribution of melanists, black hamsters occur everywhere with but very few exceptions only. In other words,

* All-Union Fur Syndicate of the USSR.

Table 1

Region	Total	Black	Percentage of black	Region	Total	Black	Percentage of black
Vinnitza	155586	1683	1.08	Nikolaev	17062	47	0.27
Voroshilovgrad	7820	3	0.04	Odessa	39700	471	1.12
Dniepropetrovsk	118448	248	0.21	Poltava	167848	13756	8.20
Zhitomir	82996	2580	3.11	Stalino	41821	1211	2.90
Zaporozhie	18079	1	0.01	Sumy	97371	498	0.51
Kamenetz-Podolsk	333002	16140	4.85	Kharkov	32956	903	2.74
Kiev	188127	3180	1.69	Chernigov	582533	19813	3.40
Kirovograd	82276	85	0.04				
				Total	1965625	60569	3.08

the main area of distribution of black hamsters in the Ukraine is practically continuous. Except for this main area, there exist four small independent «islands» where melanists are found, but their frequency is very low there (in two of them only a single black individual was trapped in each, during the period studied).

Within the limits of the area of distribution of black hamsters there exist four independent zones where the frequency of melanists is especially high (5—20 per cent and more). Nearly all the districts having an average frequency of black hamsters exceeding 5 per cent are concentrated in these zones. The first of these zones is located in the southern part of the Chernigov province and the north-western part of the Poltava province. The second lies in the south-eastern part of the Poltava province. The third is located in the western part of the Kamenetz-Podolsk province. The fourth lies in the north-western part of the Kamenetz-Podolsk province, the north-western part of the Vinnitza province and the south-western part of the Zhitomir province.

It is highly interesting that these zones are nearly everywhere outlined very sharply, as is well illustrated by Table 2.

Table 2

Zone	Strap of land 20 km broad inside the boundaries of the zone			Strap of land 20 km broad inside the boundaries of the zone		
	Number of points	Mean frequency of melanists, %	Range, %	Number of points	Mean frequency of melanists, %	Range, %
I	8	17.58	10.67—27.14	11	1.49	0.00—4.52
II	11	16.07	5.00—41.48	13	1.68	0.00—3.80
III	7	8.87	6.34—13.52	9	2.45	0.52—4.40
IV	8	7.46	5.31—10.45	10	1.97	0.00—4.85

Beyond these four zones characterized by a high frequency of melanists lies a large number of districts with a much lower frequency of black hamsters (usually below 1 per cent). No safe grouping of these districts on the basis of the frequency of melanists is possible; it is probable, however, that in the vicinity of the zone of high frequency of melanists the percentage of melanists is somewhat higher than in remote districts.

The continuity of the area of distribution of melanists and the concentration of all the districts with a high frequency of black hamsters into sharply outlined zones make it evident that the presence of black hamsters in some localities and

their absence in others, as well as the differences in the frequency of melanists within the area of their distribution, cannot be explained by random fluctuations in the reproduction and survival of the two forms of hamsters. Three other possible explanations remain.

1. It may be assumed that the black mutation of the hamster is «diffusing» from the point or points of its origin, but had not enough time as yet to spread throughout the area of distribution of the species.

2. The peculiarities of the distribution of melanists may be caused by isolation.

3. The observed picture of distribution of melanists may be determined by the fact that in some localities ecological conditions favour black hamsters and in other they favour commonly-coloured ones so that natural selection prevents the spread of the melanic mutation all over the area occupied by the species, leading to differences in the frequency of black hamsters within the area of their distribution.

A number of considerations make one believe that diffusion and isolation have played but a very insignificant part in limiting the distribution of black hamsters over the territory investigated. In the first place it should be noted that, as shown by us, the melanic mutation of the hamster is capable of spreading very rapidly under favourable conditions, this being undoubtedly connected with the high reproduction rate of the hamster and with the migrations which frequently occur in this species. Black hamsters are known in the Ukraine for about a century, and during this time they could have spread all over the hamster populations. The absence of any gradient in the frequency of melanists in those regions of the Ukraine where they exist also speaks against the free diffusion of the melanic mutation from a definite centre of origin.

It is also difficult to explain through isolation the absence of black hamsters in some parts of the Ukraine and the differences in their frequency in other parts. The data collected from hundreds of factories, as well as our field collections show that hamsters are present everywhere over the investigated territory, with the possible exception of the northern part of the Zhitomir province. Thus, no significantly large spaces exist here which are uninhabitable for hamsters. No natural barriers exist between the districts where melanists are present and those where they are absent, neither do such barriers separate the zones of high frequency of melanists from the surrounding districts where melanists are relatively rare.

Thus, only the third possible explanation remains, *viz.* the assumption that ecological conditions in some districts favour black hamsters, whereas in others these conditions prevent their survival. However, this assumption cannot be regarded as sufficiently well grounded unless it is shown that such differences in the ecological conditions of the corresponding areas actually exist, and that the boundaries separating districts with different frequencies of melanists coincide with the boundaries of such ecological zones.

In order to solve this question we undertook a careful comparison of the map of distribution of black hamsters in the Ukraine with corresponding maps of the relief, climatic, soil and geobotanical maps. We also carried out a special field investigation of the ecological conditions in the Chernigov zone of high frequency of melanists and in the neighbouring districts.

This comparison has revealed the existence of a clear and close connexion between the distribution of black hamsters and the type of vegetation. Melanic hamsters are found in the Ukraine almost exclusively in the forest-steppe zone, and the boundaries of this zone coincide fairly well with the boundaries of the area of distribution of melanists. Within the forest-steppe zone black hamsters are most frequent in lowland, moist places.

Thus, it seems most probable that natural selection is the chief cause determining the peculiarities of the geographical distribution of black hamsters in the Ukraine.

It is interesting to note that a high positive correlation exists between the frequency of melanists and the density of the hamster population. In order to

illustrate this, Table 3 presents the relative density of the hamster population in two regions of the Ukraine, characterized by a high frequency of melanists, and in two regions where melanists are rare (data for 1935—1938). A similar correlation exists in other regions, as well as in the districts whithin each region. This phenomenon may be connected with the observation of Petzsch [3] that the frequency of melanists increases during the years of mass reproduction of hamsters.

Table 3

Region	Number of hamsters trapped *per annum* per km²	Percentage of melanists
Kamenetz-Podolsk	3.35	4.91
Poltava	1.07	7.80
Voroshilovgrad	0.04	0.04
Zaporozhie	0.10	0.01

In respect to the four small «islands» of distribution of melanists, located beyond the main area of their distribution, our data show conclusively that in two of them the melanic mutation arose independently, and make the same highly probable for the third.

Received
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Comptes Rendus (Doklady) de l'Académie des Sciences de l'URSS
1945. Volume XLVII, № 8

ZOOLOGY

EFFECT OF SUN RADIATION UPON BODY TEMPERATURE OF CERTAIN LITTORAL ANIMALS

By I. D. STRELNIKOV

(Communicated by I. I. Schmalhausen, Member of the Academy, 30. VII. 1944)

Littoral animals that live above sea-level are constantly exposed to the influence of the various agents of the air medium, *viz.* temperature, sun radiation and scattered light, wind and humidity. Such animals as occupy the zone between high-and low-water mark, are subject to the influence of the air medium during the ebb-tide only.

When in water, the body temperature of invertebrates is equal to that of the water. Aquatic animals experience no thermal influence of sun radiation when they are in water. The surface layers of water absorb the heat. Owing to the high thermal conductivity of water, the heat is immediately transferred from the organism to the surrounding medium, *viz.* water.

I undertook to investigate the life conditions of littoral animals. For this purpose measurements of temperature were made for a number of molluscs and crustaceans, as well as air temperature measurements, and records of the effect of sun radiation upon the animals under observation. The work was carried out at the Karadag biological station situated at the Black Sea shore.

The body temperature of the animals and the temperature of the surface of the cliffs were measured with the aid of Zeiss' loop galvanometer (1). The intensity of sun radiation was measured by means of Kalitin's pyranometer and Zeiss' galvanometer.

The records given in Table 1 show that the body temperature of the *Chtamalus stellatus* exposed to the influence of sun rays, fluctuates within the range of 30—38°C. The body temperature of the animals kept in shade is found to be considerably lower, *viz.* 24.4°C. The body temperature of the animals kept in water was found to be equal to that of the water, *i.e.* 22.8°C.

The cliffs to which *Ch. stellatus* were attached, were exposed to sun radiation; the temperature of the surface attained 39.6 C in the open, and 38°C in the fissures of the cliff.

The body temperature of *Ch. stellatus* is dependent on the sun radiation, the temperature of the cliff to which the animals are attached, and on the temperature of the air. Even when kept in the shade the animals show a body temperature lower than that of the surrounding air medium; this is due to the effect of the rock which reduces the body temperature.

When the animals are in water, the temperature is evenly distributed all over their bodies. When above water mark, exposed to the effect of the heat of direct sun rays, or of heat transmitted through the heated rock, or else, to the cooling effect of air or wind, or to that of the cold cliff, thermal currents are formed in the organism of the crustaceans. It should be assumed that these currents play an essential part in the physiological processes which take place in the organism of those animals.

Table 1

Effect of Sun Radiation upon Body Temperature of *Chtamalus stellatus* Poli (sun radiation=0.78 gcal/cm² min.)

Day and hour (a. m.)	Air temperature, °C	Body temperature, °C	Note
23.IX			
10 h. 10 m.	29.2	37.7	At a cliff 30 cm above water level, side illumination at an angle of 90°
10 h. 15 m.	»	36.9	Side illumination
10 h. 16 m.	»	34.2	Base in the shade, in a fissure, side illumination
10 h. 18 m.	27.6	37.4	Side illumination
10 h. 19 m.	»	32.0	In a narrow fissure of a cliff, moist
10 h. 20 m.	»	32.0	At a height of 20 cm, at the mark between dry and moist cliff
10 h. 21 m.	»	32.3	Base of animal shaded, upper part of body exposed to light, 20 cm above water mark
10 h. 22 m.	25.3	34.7	Illuminated at an angle of 80°, 40 cm above water mark
10 h. 24 m.	»	34.6	At a height of 25 cm, illuminated from the side and from above
10 h. 25 m.	25.6	34.0	At a height of 60 cm, total exposure to light
10 h. 28 m.	»	22.8	Immersed in water, in the very surface layer
10 h. 31 m.	»	30.5	Illuminated at an angle of 30°
10 h. 32 m.	29.4	34.4	Side illumination
10 h. 33 m.	»	33.4	Illuminated at an angle of 90°; in shade until 2 min. before $t°$ measurement
10 h. 34 m.	»	38.0	At a cliff illuminated at an angle of 45°
10 h. 37 m.	29.	37.9	Near the previous animal, at a height of 40 cm
10 h. 38 m.	»	37.1	At a height of 25 cm, illuminated at an angle of 45°
10 h. 40 m.	»	36.9	Side illumination
10 h. 41 m.	»	36.7	At a height of 25 cm, illuminated at an angle of 15°
10 h. 45 m.	»	24.4	In the shade, at a height of 80 cm

Moistening of the body surface with water or even slight spraying, results in a cooling and reduction of body temperature.

Owing to their solid calcareous shell and the tightness of the separate elements of their external skeleton, *Ch. stellatus* are prevented from drying even if subjected to high temperatures caused by sun rays, to which they are exposed when outside the water.

On account of a possible overheating due to sun rays, *Ch. stellatus* prefers such places as are exposed to heavy breakers; however, the animals settle largely on the shaded side of the rocks, opposite to that exposed to the sea and the sun (²).

Table 2

Body Temperature of *Pachygrapsus marmoratus* as Affected by Sun Radiation (sun radiation 0.9 gcal/cm² min.)

Hour (a. m.)	Air temperature, °C	Body temperature, °C	Note
11 h. 0 m.	17.6	27.3	Small crab, 1 cm across
11 h. 30 sec.	»	22.2	Same crab; $t°$ somewhat reduced owing to evaporation
11 h. 2 m.	»	28.0	Same crab, after exposure to sun rays
11 h. 3 m.	»	27.7	Small crab from a pool on the cliff
11 h. 4 m.	»	26.4	Adult crab from the same pool, water temperature 26.4°

These measurements demonstrate that body temperature of crabbs when outside the water, or in small water basins on cliffs, is higher than the temperature of sea water in the surface layer (22.1°), the difference amounting to 5°C.

Table 3

Body Temperature of the Molluscs *Patella pontica* and *Mytilus galloprovincialis* as Affected by Sun Radiation (sun radiation 0.95 gcal/cm² min.)

Hour	Air temperature, °C	Body temperature, °C	Note
11 h. 57 m. a. m.	27.6	38.5	*t*° of *P. pontica* outside the water
12 h. 1 m. p. m.	»	38.8	Same
12 h. 3 m. » »	»	37.9	»
12 h. 4 m. » »	»	26.4	Same, reached by water (of *t*° 25.1° C)
12 h. 6 m. » »	»	40.1	Same, outside water
12 h. 12 m. » »	»	40.0	Same
12 h. 13 m. » »	»	33.3	Half body in shade
11 h. 6 m. a. m.	»	26.4	Body temperature of *P. pontica* in a pool on the cliff

During the time they are outside the water, littoral molluscs are exposed to the effect of sun radiation and air heat. As may be seen from my measurements, the temperature of *Ch. stellatus* and *P. pontica* rises to 40°C outside the water, whereas in the surface layer of water the same molluscs show a body temperature similar to that of the water, *i. e.* 22.8—23.1°C.

Littoral molluscs and crustaceans are able to resist to the abrupt changes in ecological conditions and to high temperatures due to sun rays, owing to the special morphological adaptations which protect these animals from drying up. These adaptations consist in the shells and in the tight fitting of the separate parts of the carapace, respectively. As well-known, *Ch. stellatus* is able to live outside the water for several days. According to my observations, *Ch. stellatus* caught in the Barents Sea could live up to 7 days without water in the laboratory.

The life in a littoral zone between water and air, the alternation of the two media, which differ sharply with respect to such ecological agents as sun radiation, temperature and humidity, creates conditions favourable for variation and formation of species.

It is in the littoral zone that had proceeded the gradual transition from the aquatic mode of life to the life on dry land, which was the most important stage in animal evolution. Like in those remote geological epochs when aquatic animals gradually abandoned the sea to live upon dry land, recent littoral animals, passing from the water medium characterized by relatively uniform conditions to the air zone, with alternating and variable conditions both in time and space, underwent morphological changes and adapted themselves to their new living conditions. Within the littoral zone the formation of new species proceeded at a higher rate than in the water medium, at the depth of seas and oceans.

Received
30. VII. 1944.

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